According To National Curriculum

A Text book of

PHYSICS

FOR CLASS IX-X

SINDH TEXTBOOK BOARD, JAMSHORO.

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PREFACE

Many developments have taken place since the publication of the existing Physics text book, published by Sindh Textbook Board, Jamshoro. Keeping in view the advances made in Physics, the Curriculum Wing of the Ministry of Education, Government of Pakistan Islamabad, revised the existing syllabus after incorporating the latest developments in the field of Physics.

This book is written for the students of Class IX-Xstrictly in accordance with approved syllabus published by the National Bureau of Curriculum, Ministry of Education, Government of Pakistan, Islamabad under the sponsorship of Sindh Textbook Board, Jamshoro. This is an official Physics Textbook for Class IX-Xin Pakistan, in general, and for the entire province of Sindh in particular.

The purpose and objectives of this text book is to introduce the students to the fundamental laws and principles of Physics with their applications. The language used throughout the book is simple, straight forward, easily understandable and is of the level of comprehension of the students of secondary classes.

The compilation of matter for the said book was a cumbersome job, but even more difficult was to maintain the sequence and continuity of language of different styles because the topics selected were written by different authors. We have endeavoured our level best to maintain the fluency of language so that a reader should not find any difficulty in understanding the subject.

Numerous solved examples have been included to illustrate the application of various theories. Some questions and a number of unsolved problems are also incorporated at the end of each chapter. However, this should not prevent the teachers and the students from using other sources to improve their problem solving skills. The authors have unsparingly consulted various pertinent references and text books related to the contents.

We are thankful to the authorities of Sindh Textbook Board, Jamshoro, specially Prof. Imtiaz Ahmed, Mr. Yousuf Ahmed Shaikh and Mr. Danish, who extended full co-operation and possible help during the preparation of the manuscript. We are deeply grateful to the members of National Review Committee of Physics for their concrete suggestions, valuable comments and constructive ideas during the review of an early manuscript of this text book.

Any constructive criticism and healthy suggestions, for improvement of the book, from the teachers, students and public in general will be gratefully acknowledged for future edition of the present volume.

Prof. Muhammad Idrees Khan
Prof. Imtiaz Ahmed Khan

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NUCLEAR PHYSICS

بسم الله الرحمن الرحيم

CHAPTER – 1

INTRODUCTION

LEARNING OBJECTIVES:

- What is Physics.
- Importance of Physics in daily life.
- Teaching of Islam and Science.
- Contribution of Muslim and Pakistani Scientists.

1.1 WHAT IS PHYSICS?

Almighty Allah created this universe billions of years ago with a single word "be" and at once it came into being. He inducted several principles and laws in it to sustain its function. Now from the day of their creation every particle of the universe is following these laws. These laws are known as "THE LAWS OF NATURE". Apparently they are hidden from the eyes of mankind and are mysteries for him. Our solar system including our earth is a small component of this immense universe. This is composed of matter and hence is a physical world.

Human being is the best creature of Allah in this world and has been endowed with many qualities. One of them is to unfold and discover the Laws of Nature. He has curiosity in his nature. When he looks around, he observes that innumerable number of events are taking place. He sees that clouds are floating in the sky rainfalls from them. Thunder occurs and lightening flashes in the clouds. Several tiny plants and beautiful grass appear on the surface of dry land. He observes that there are many objects that catch fire while others do not. Heat burns things and destroys their identity. Some matter appears in solid state, some in liquid and other in gaseous state. These and many other observations raise several questions in his mind. He asks where and how clouds are formed. How lightening flashes in the clouds? Why some objects catch fire and others do not? What is the nature of heat? How is it produced? How matter

appears in different states? Such and many similar questions excite him to find the answers of these questions. Physics a branch of science which helps him to get the answers of his queries.

The word "Physics" comes from a Greek word "Physikos" meaning "natural." Now it is defined as a science which deals with the interaction of matter and energy. The physicists have divided the subject of physics into various branches; each branch has acquired a title of a full subject nowadays. Main branches of physics are given below:

1. **Mechanics:** It deals with the motion of objects with or without reference of force.

2. **Electricity:** It is concerned with the phenomena and effects related to electric charges.

3. **Electromagnetism:** It deals with observations, principles, laws and methods that relate electricity and magnetism.

4. **Solid-State physics:** It is concerned with the structure and properties of solid materials.

5. **Atomic physics:** It is concerned with the structure and properties of the atom.

6. **Nuclear physics:** This branch is concerned with the structure, properties and reactions of the nuclei of atoms.

7. **Plasma physics:** It is concerned with the properties of highly ionized atoms forming a mixture of bare nuclei (nuclei without electrons) and electrons.

8. **Bio-physics:** It is concerned with the application of physical methods and explanation to bio-physical systems and structures.

9. **Astro physics:** It is concerned with the study of the physics of astronomical bodies.

In addition to these branches also there are other branches, say low temperature physics, particle physics, optics, etc.

1.2 IMPORTANCE OF PHYSICS IN DAILY LIFE:

In our daily life we use many things as a routine. But we seldom think how they are made. We use several electric appliances in our homes. For example, electric fan, electric bulb, refrigerator, airconditioner, juicer, grinder, etc. They all use electric power. We use buses, cars, railway carriages, aeroplanes etc., for long distances. All of these are run by engines. Engines of these vehicles ar manufactured on the principles of thermodynamics. We listen to rad

programmes. We see the events far away from us on the television screen and enjoy various programmes telecast from the T.V. station. In all electronic devices and appliances solid state physics plays a vital role. Laser technology is widely used in defence system, metallurgy, medical science & astronomy which has its roots in atomic physics. Many countries including Pakistan, utilize nuclear energy to produce electric power. Production of nuclear energy is the area of nuclear physics.

We can see that in every walk of life physics is involved in one way or the other. Thus physics is playing a vital role in framing our life style and influencing our way of thinking.

Scientific Method:

In a scientific work the most important thing is observation. Observations are taken very carefully and systematically. In this way all possible informations are gathered about a physical phenomenon under investigation. Keeping these observations in mind a scientist makes some hypothesis or postulate to explain the phenomenon. A hypothesis is a notion of a scientist on which he tries to explain the phenomenon. He designs a theory on the basis of hypothesis and gives an elaborated explanation about the phenomenon. The theory is checked by experiments. If theoretical and experimental results agree with each other, the theory is regarded correct otherwise it is discarded. This is one way of investigation of a problem in physics. Some times a scientist presents a theory on the basis of rational reasoning and predicts some phenomena to take place under certain conditions. Experiments are designed to test this theory. Again if predicted results are obtained, the theory is supposed to be correct and then this theory becomes a law. The law is such a statement regarding the behaviour of nature which explains the observations and experiments of the past and can predict about other aspects of nature.

Newton made a hypothesis about the nature of light. According to him light consists of extremely small particles moving with a very high velocity. He named these particles as corpuscles. Based on this hypothesis he developed a theory, called corpuscular theory of light, which could explain the phenomenon of reflection and refraction of light and the formation of shadows which were experimental facts. Thus theory was accepted because it could explain a few properties of light as mentioned above. According to corpuscular theory of light, light travels faster in a denser medium (water) than rarer medium (air) whereas, according to wave theory light travels faster in rarer medium than a denser medium. The phenomena of interference, diffraction and polarization of light can not be explain on the basis of corpuscular theory of light. These phenomena can be explained successfully with help of wave theory of light. Foucault determined speed of light in water experimentally which showed that light

travels slower in water than air. This result was against the corpuscular theory of light hence the theory was discarded and wave theory was accepted. Later on phenomena of photo electric effect and compton effect were observed which could not be explained with the help of wave theory of light. Einstein presented another theory regarding the nature of light which is known as quantum theory of light. According to this theory light has dual nature; some times it behaves like a wave and some times like a particle. It consists of packets of energy, called photons. All the phenomena regarding the nature of light are explained with the help of the quantum theory of light.

Errors and accuracy:

In an experiment certain errors may arise due to experimenter, or to the instrument used or due to both. We divide these errors into three categories, namely personal error, systematic error and random error.

(a) **Personal error:** This error arises due to tendency of a person who takes readings in favour of a particular reading. Such procedure produces errors in taking observations. This type of error can be removed by avoiding this bias.

(b) **Systematic error:** This type of error arises due to a fault in the measuring instrument. This is called the zero error of the instrument. This error can be overcome by adopting the relevant zero error correction.

(c) **Random error:** This type of error may arise due to external conditions which are at a certain time beyond the control of the experimenter. This error can be minimized by maintaining certain conditions in the laboratory. For example, changes in temperature, humidity and voltage etc. may be controlled.

During the measurement of a quantity. The maximum possible error of an instrument is defined as half the distance between two consecutive divisions. Suppose a measuring rod is marked in millimetres. The likely error in the measurement on this scale would be 0.5 mm at one end. If a rod to be measured is placed along the scale the likely error in the measurement will be 0.5 mm+ 0.5 mm = 1 mm or 0.5 cm. It is because at both ends the likely error is 0.5 mm. This error may be positive or negative, because in the measurement both possibilities are there. If the rod is measured 16.2 cm in length, then its numerical value shoud be written as (16.2 ± 0.5) cm. It means that likely error in measurement is 0.1 cm. This error is usually expressed in percent, hence the relative possible error in the above example is $\frac{0.1}{16.2} = 0.006$ and percentage. error, $\frac{0.1}{16.2} \times \frac{100}{100} = 0.6\%$.

Consider another example of measurement in which a thermometer is used. It is calibrated in 0.2°C division. In an experiment initial and final temperatures are recorded as 20.2°C and 25.4°C. At each reading the likely error is ±0.1°C. Therefore the likely error in rise in temperature (25.4°C – 20.2°C = 5.2°C), would be 0.1°C + 0.1°C = 0.2°C. The percentage error will be $\frac{0.2}{5.2} \times 100 = 3.8\%$.

Usually in an experiment we use a formula to find the final result. If a formula contains one or more quantities raised to some power, say, volume of a sphere is given by the formula $V = \frac{4}{3}\pi r^3$, where r stands for the radius of the sphere, then the error in the result increases due to the quantiy multiplied by the power to which the quantity is raised in the original formula. In the above example,

$$V = \frac{4}{3}\pi r^3 = \frac{4}{3}\pi\left(\frac{d}{2}\right)^3 = \frac{4}{3}\pi\frac{d^3}{8}$$

$$= \frac{\pi d^3}{6} \text{ (here d stands for diameter)}$$

Percentage error in V = 3 × %age error in d.

since the diameter is raised to the power three.

If two or more quantities are measured in an experiment, then the error in the final result will be the sum of errors of individual quantity. For example the volume of a box is given as,

$$V = l \times b \times h$$

then

Percentage error in volume = %age error in length +%age error in breadth +%age error in height

If quotient is involved in a formula, say, density of a material is given as,

$$\rho = \frac{\text{Mass}}{\text{Volume}}$$

then percentage error in the result would again be the sum of percentage error in individual quantity.

Thus percentage error in density = % error of mass + % error of volume.

Some times we take several readings of the same quantity to improve the accuracy in the result. Suppose six readings are taken on a screw gauge (micrometer) for the diameter of a wire. The diameter of a wire expressed in mm is

$$1.23,\ 1.22,\ 1.25,\ 1.25,\ 1.23,\ 1.23$$

The mean of these readings is

$$\bar{x} = \frac{1.23 + 1.22 + 1.25 + 1.25 + 1.23 + 1.23}{6}$$

$$= 1.24 \text{ mm}$$

Now we find the deviation for each reading from the mean value irrespective of its sign.

Thus,

actual reading	1.23, 1.22, 1.25, 1.25, 1.23, 1.23
mean reading	1.24, 1.24, 1.24, 1.24, 1.24, 1.24
deviation	0.01, 0.02, 0.01, 0.01, 0.01, 0.01

Now we find the mean of deviations.

$$\overline{D} = \frac{0.01 + 0.02 + 0.01 + 0.01 + 0.01 + 0.01}{6}$$

$$= 0.01 \text{ mm}$$

This is the likely error in the mean reading 1.24 mm. Now the mean reading along with its likely error will be written as (1.24 ± 0.01) mm.

1.3 TEACHINGS OF ISLAM AND SCIENCE:

In an authentic hadith Saheeh Bukhari Sharif the mode of the first revelation is narrated when the Holy Prophet (P.B.U.H.) was in the cave of Hira.

"Read: In the name of thy Lord Who createth."
"Createth man from a clot.:
"Read: And it is thy Lord the most Bountiful."
"Who teacheth by pen, Teacheth man that which he knew not."

(Surah: Al-Alaq – Aayat 1-5)

It is obvious from the first revelation that the reading in Islam is the most important feature of life. In the Holy Quran the knowledge of names of things

is related to the very first man, Adam (may Allah place his soul in peace and tranquillity). Superiority of Adam to angels is due to knowledge of things.

> "And He taught Adam all the names, then showed them to the angels, saying:
> Inform me of the names of these if ye are truthful."
> "They said: Be glorified! We have no knowledge saving that which Thou hast taught us. Lo! Thou, only Thou, art the Knower, the Wise."
> "He said: O Adam! Inform them of their names, and when he had informed them of their names, He said: Did I not tell you that I known the secret of the heavens and the earth? And I know that which ye disclose and which ye hide".
>
> *(Surah: Al-Baqarah – Aayat 31-33)*

These verses tell us that the first man, Adam, created on the surface of the earth had knowledge of things and was not an ignorant person. It means that our very first ancestors were blessed with knowledge of things by Allah Himself. Adam (may Allah place his soul in peace and tranquillity) was not only the first man on the earth but also the first prophet or the messenger of Allah. Therefore teachings of Islam from its very beginning is to acquire knowledge of things.

One hadith narrates: "Quest knowledge from cradle to grave". Our Holy Prophet (peace be upon him) said: "Gain knowledge if you had to travel even to China". In the days of Prophet China was supposed a far remote land from Arabia, as well as learning seat of knowledge. In another hadith the Holy Prophet (peace be upon him) said: "To gain knowledge is an obligation to every muslim man and woman". These and many other hadiths indicate the importance of learning and knowledge in Islam. In the light of teachings of Islam we cannot ignore learning and gaining knowledge from all quarters of life. Of course, knowledge is the legacy of a believer.

It was the impression of teachings of Islam that a great number of muslim scholars indulged in investigation of natural phenomena and testing old theories practically. In this way they opened the door of practical science. Now we state a few renowned scholars of muslim world and their prominent work in the field of science.

1.4 CONTRIBUTION OF MUSLIM AND PAKISTANI SCIENTISTS:

In Abbasid period Baghdad was a great centre of learning and knowledge. Scientists and intellectuals from all over the world came here to quench their thirst for knowledge. Khalifa was a lover of knowledge and he encouraged learners and scholars for their remarkable achievements in the field of science.

Abu Ali Hassan Ibn-al-Haitham (965-1039 A.D)


Abu Ali Hassan Ibn-al-Haitham

Abu Ali Hassan Ibn-al-Haitham was born in Basra, a renowned city in Iraq. He was a great scholar of physics, mathematics, engineering, astronomy and medicine. He wrote many books on various subjects. His book on optics, Kitabul Manazir, gained great reputation among scholars. In this book he criticised the theory about the nature of light, presented by the great philosophers of the past such as Ptolemy and Euclid. It was his courage that he refuted this theory on practical ground.

According to this theory firstly light enters the eye then emits in the form of rays from the eye. These rays fall on the object and make it visible. But Al-Haitham pointed out that when light is incident on an object some of it is reflected back and enters the eye. As a result the object becomes visible to the eye. He discovered the nature of light and told that it is a kind of energy. He gave formal definition of ray of light. He invented a pin hole camera and with this he obtained an image of sun eclipse. He gave two laws of reflection and carried out research on the formation of images by spherical mirrors. In his book the chapter on human eye is still a remarkable collection of information on this topic. Another important principle he discovered is that a ray of light passing through a medium takes a path which takes the least time to traverse.

Yaqub Ibne Ishaq-Alkindi (800-873 A.D.)

Yaqub Kindi (800 A.D) was born in Basra. He produced several research monographs on meteorology, specific gravity and on tides. His most important work was in sound and optics. He explained musical notes on scientific ground and discovered a method to express the notes in terms of frequencies and used it to fix the order of various notes according to their frequency. He also discussed the nature of sound. He worked in geometrical optics which was translated in Latin. He was a man of diverse nature and he did extensive work on different fields of science and promoted new thoughts in many disciplines of learning.


Yaqub Ibne Ishaq-Alkindi

Abu Rehan Muhammad bin Ahmed Al-Beruni

Al-Beruni

Al-Beruni was born in a small town Berun in present Afghanistan in 973 A.D. He was a great scholar of his time. He wrote more than one hundred and fifty books on various subjects, such as, mathematics, physics, cosmology, geography, history, culture and civilization, archaeology, comparative religions, geology, chemistry and biology etc. He explained every subject from a new angle and refuted many theories about the nature of things. For example, in his time it was supposed and was taken as a fact that earth was a flat body. But, not only rejected this notion but practically proved that earth is round and not flat. He also measured the circumference of earth. He discussed the movement of the sun and the moon, the phases of the moon and also the movement of the planets known in those days. He gave the method of determining the longitude and latitude of a place. He also found densities of various metals. While he was in Pind Dadan Khan, a village in Pakistan, he determined the circumference of the earth by his own invented method. He told that the earth is floating in space just like a grape floats in water. He rejected the notion that earth was stationary body in the universe. It was he who disclosed that Indus Valley was once the basin of an ocean. It was gradually filled up by mud. Now it has been confirmed by modern geologists.

Dr. Abdus Salam

Dr. Abdus Salam was born in Jhang, a small city in Pakistan, in 1926. He was very intelligent from his very childhood. He passed every examination with distinction. Because of his good performance in education he was awarded scholarship for higher studies in U.K. He came back in Pakistan in 1950 but he again went to U.K. to promote his research work.

Dr. Abdus Salam

He was awarded Nobel Prize in physics in 1979 for his work on Grand Unification Theory (GUT). He established International Centre for Theoretical Physics at Trieste, Italy where scientists from the developing countries are provided opportunities to augment their research work in which they are engaged in their own countries by joining with elites of physics. Of course, he was an asset of Pakistan and will remain in the hearts of learning circle.

Dr. Abdul Qadeer Khan

Dr. Abdul Qadeer Khan

He was born on 1st April, 1936 at Bhopal in India. He obtained M.Sc. Metallergy degree from Holland. He was selected as research Assistant in the same University. He obtained Ph.D degree from the University of Leaven Belgium. He worked as an expert at Urenco Enrichment Plant in Holland as a Joint Venture of the Government of Holland. When Dr. Abdul Qadeer Khan imbued with the supreme spirit of patriotism, he returned to Pakistan to serve his motherland. To honour him, the former Engineering Research Laboratories has now been named as A.Q. Khan Research Laboratories. He has been awarded Hilal-i-Imtiaz by the Government of Pakistan. He contributed in making Pakistan a nuclear state.

SUMMARY

1. **Physics:** The branch of science which deals with the study of properties of matter and energy along with their mutual interaction.

2. **Branches of Physics:** Mechanics, Light, Sound, Heat & Thermodynamics, Electromagnetism, Atomic Physics, Nuclear Physics, Plasma Physics, Solid State Physics, Molecular Physics, Astro-Physics, etc.

3. **Scientific Method:** It is a specific method used for the search of the truth. It consists of the following steps:

   (i) Observation (ii) Hypothesis (iii) Experiment
   (iv) Theory (v) Prediction (vi) Law

4. **Islamic Education and Science:** In the Holy Quran, The Almighty Allah urges the Muslims to study nature, to act with sobriety and to benefit from scientific knowledge thoughtfully.

5. **Muslim Scientists:** Muslim scientists emphasized the importance of observations and experimentation in the study of science.

   The names of Al-Beruni, Ibn-ul-Haitham and Al-Kindi are significiently renowned and their light of knowledge is beneficial to scholars even today.

6. **Pakistani Scientist:** The only Nobel Prize Holder from Pakistan is Dr. Abdus-Salam.

QUESTIONS

1.1 Write briefly the answers to the following questions.

(a) What is Physics? Name a few branches of Physics?
(b) What is the contribution of Al-Haitham in the field of Physics?
(c) What is the contribution of Dr. Abdus Salam in the field of Physics?
(d) In what way Al-Beruni was versatile scientist?
(e) State briefly the significant work of Dr. Abdul Qadeer Khan in the field of nuclear science.

1.2 Fill in the blanks.

(i) Physics is that branch of science which deals with the study of the properties of matter and ____________ and interaction between them.
(ii) Science is the common ___________ of all mankind.
(iii) Knowledge is the _________ of a believer.
(iv) Laser is the field of ____________ Physics.
(v) Al-Beruni gave the method of determining the _________ and __________ of a place.

1.3 Given below are a few possible answers to each statement; identify the correct one.

(i) Ibn-al-Haitham contributed toward _________ Physics.
(a) Nuclear. (b) Oceanographic.
(c) Optical. (d) Thermal.

(ii) The name of Muslim scientist who was born in Basra and made several discoveries on music was ___________.
(a) Al-Beruni. (b) Al-Kindi.
(c) Al-Khwarzmi. (d) Nasir-ud-Din Tusi.

(iii) Dr. Abdus Salam was awarded Nobel Prize for his work on ____________.
(a) Electronics. (b) Radiation.
(c) Grand Unification Theory. (d) Gravitatian.

1.4 Pick out true and false.

(i) The first revelation came on the Holy Prophet (P.B.U.H.) was about the creation of heaven.
(ii) Pin Hole camera was invented by Ibn-al-Haitham.
(iii) Al-Beruni proposed that the Indus Valley was the basin of an ancient ocean that has gradually been filled up by mud.
(iv) To gain knowledge is an obligation for man only. This is a basic principle of Islam.
(v) The behaviour of atoms in the isolated state is studied under the branch of plasma physics.

CHAPTER - 2

MEASUREMENT

LEARNING OBJECTIVES:

- Physical quantities.
- Derived units.
- Significant figures and accuracy.
- Measuring instruments.

INTRODUCTION

Measurement is the common practice of every day life. This routine work starts from morning till late hours in the night. Every morning a milk man comes and gives a measured volume of milk to the house hold with the help of his measuring cylinder graduated in halves of a decilitre. If one goes to a shop to purchase sugar, the shopkeeper will weigh the required amount of sugar by his common balance and will handover it to the purchaser. To purchase cloth one goes to the shop and the shopkeeper will measure the required length of the cloth by his metre scale graduated in centimetres.

From these examples it is clear that to measure volume, mass and length we use measuring cylinder, common balance and the metre scale respectively as the measuring devices or instruments.

Similarly if a college student goes to his laboratory to perform experiments, he may be asked by the laboratory incharge to find the thickness of a thin glass slab, the radius of curvature of a lens or density of a metal block so on and so forth. The student will use the measuring devices (generally called apparatus) suiting his experiments.

Thus the physical quantities like length, mass, volume, time, current, voltage and pressure etc are measured by the instrument (measuring devices) of the relevant type. The accuracy for measurement depends on the quality of

the instrument used. For example to measure the length of the rod in the laboratory we use the scale graduated in centimetres. The least measurement that this scale can make correctly is 1cm (one centimetre). This least measurement is called the least count of the scale.

In order to measure the small lengths like the thickness of a thin glass slab, the radius of thin wires used in laboratories etc we use vernier callipers or micrometre screw guage. They can measure correctly upto 0.1mm or 0.01cm and 0.01mm or 0.001cm respectively. They are called the least counts of the respective measuring instruments.

2.1 PHYSICAL QUANTITIES

Almighty Allah has created every thing in this world with the best estimation. From a huge celestial body to a sub-particle of an atom every one is working in its domain under certain principle which has been assigned for it by Allah. Man has to discover these principles. A physicist is largely involved in discovering the law of nature through precise measurement of various physical quantities. Scientists have divided all physical quantities into two groups, (a) fundamental quantities, (b) derived quantities. In physics length, mass and time are supposed to be the main fundamental quantities since scientists all over the world have recognized that all physical quantities in mechanics can be expressed in terms of these fundamental quantities. To measure a physical quantity we need a certain unit. So the units to express fundamental quantities are known as FUNDAMENTAL UNITS, and the units used to express other physical quantities that are derived from fundamental units are called DERIVED UNITS. But what is meant by the word "unit"? A unit is fixed by definition and a STANDARD is an embodiment of a unit under certain physical conditions. Measurement means the comparison of an unknown quantity with a standard to see how many times it is big or small as compared to the standard.

A set of fundamental and derived units is known as a SYSTEM OF UNITS. There are three systems of units being used in scientific work. In the CGS system, centimetre, gram and second are the fundamental units for length, mass and time respectively. International system of units is abbreviated as SI from the French "Le System International d' unités". In this system seven quantities have been accepted as fundamental quantities. They are length, mass, time, electric current, amount of substance, thermodynamic temperature and luminous intensity. Units of these quantities are metre for length, kilogram for mass, second for time, ampere for electric current, mole for amount of substance, kelvin for thermodynamic temperature and candela for luminous intensity.

SI system of units is convenient for scientific work and provides a simple method for calculations.

British Engineering System is an old system. In this system fundamental quantities are length, force and time. Note that in this system mass is a derived quantity and its unit is derived from the force unit i.e., pound, and is called slug. (1 slug = 32.17 lb mass = 14.59 kg). The basic unit for length, force and time are foot, pound and second respectively.

Now-a-days SI units are used throughout the world, so we define base units of this system.

Standard of Length - metre

The current definition of meter, which is defined as:

"The metre is the length of the path travelled by light in vacuum during a time interval of $\frac{1}{299{,}792{,}458}$ of a second".

It is also defined as a distance between two marks engraved on an irridium-platinium alloy bar kept at international bureau of weights and measures near Paris.

Standard of Time - Second

In past the spin motion of the earth about its axis and its orbital motion around the sun have been used to define a second. In October 1967 the time standard again redefined in terms of an atomic clock. The atomic clock uses cesium of mass number 133. According to this clock, "a second is defined to be exactly equal to the time interval of 9,192,631,770 vibrations of atoms of cesium 133". It may be possible that two cesium clocks running over a period of 300,000 years will differ by 1 second only. Another clock has been invented which uses hydrogen and known as hydrogen maser clock. This clock has achieved a precision of 1 second in 30,000,000 years. It means that if two hydrogen maser clocks were started simultaneously and they were allowed to run over a period of 30,000,000 years then the expected difference between their time would only be 1 second. The advantage of this method is that the cesium clock or hydrogen maser clock can be constructed in a conventional laboratory and the standard of time, second, can be checked.

It is to be noted that no matter how the standard is defined, the method always involve some sort of periodic motion.

Standard of Mass – Kilogram

One kilogram is the mass of a platinum-iridium alloy cylinder which is kept at the International Bureau of Weights and Measures in Sèvres, near Paris. Replica of this internationally agreed standard of mass have been sent to several countries.

Multiples and sub-multiples of metre, second and kilogram are given below:

1 km (kilometer) = 1000 m = 10^3m

1 cm (centimeter) = $\frac{1}{100}$ m = 10^{-2}m

1 mm (millimeter) = $\frac{1}{1000}$ m = 10^{-3}m

1 μm (micrometer) = $\frac{1}{1,000,000}$ m = 10^{-6}m

1 nm(nanometer) = $\frac{1}{1,000,000,000}$ = m = 10^{-9}m

1 min (minute) = 60 s

1 h (hour) = 60×60 s = 3600 s

1 ms (millisecond) = $\frac{1}{1000}$ s = 10^{-3}s

1 μs (microsecond) = $\frac{1}{1,000,000}$ s = 10^{-6}s

1 ns (nanosecond) = $\frac{1}{1,000,000,000}$ s = 10^{-9}s

1 g (gram) = $\frac{1}{1000}$ kg = 10^{-3}kg

1 mg (milligram) = $\frac{1}{1,000,000}$ kg = 10^{-6}kg

1 μg (microgram) = $\frac{1}{1,000,000,000}$ kg = 10^{-9}kg

c. **Hydraulic Press:**

The working of hydraulic press is similar to that of a hydraulic lift. In a hydraulic press the piston of large cross sectional area used to lift an object is provided with a rigid roof over it. When the piston B of the large area moves upward, it compresses only material placed between the piston and the roof. The hydraulic press is commonly used for compressing soft materials like cotton into compact bales. Refer Fig (10.14)

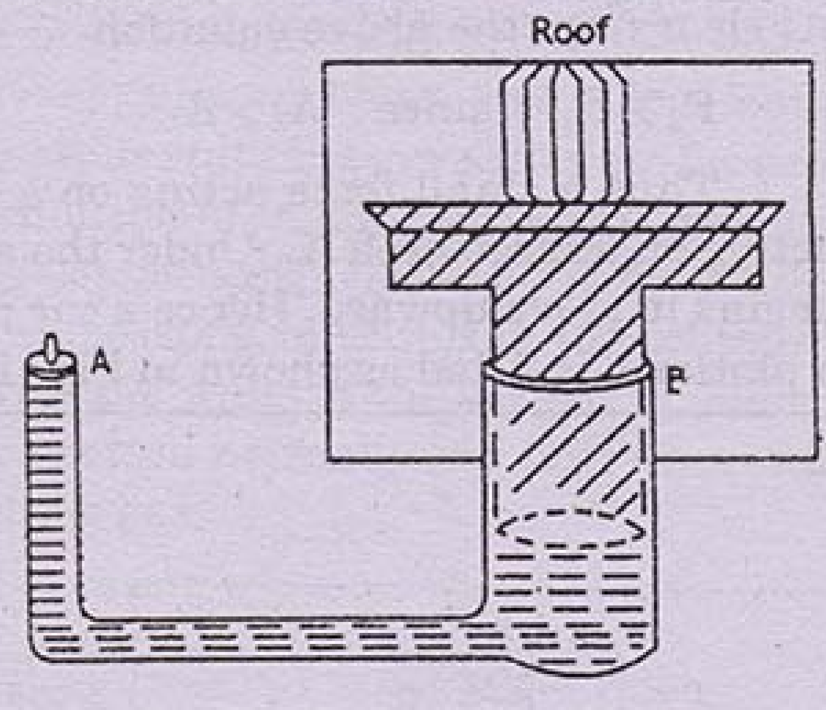


Fig. 10.9: Hydraulic Press

10.10 ARCHIMEDES PRINCIPLE

Perhaps you might have seen the ships made of iron with their masses in tons floating on the surface of water in oceans and seas and the needles also made of iron with their masses in grams and milligrams sinking in water. It is very surprising that a heavy body floats and a light body sinks. This means that for floating or sinking of a body it is not its mass which is important. It is something else which is important. The secret of the so called unlogical behaviour of floating and sinking bodies was made out in 300 B.C by a Greek scientist Archimedes. He discovered a principle regarding the floating or sinking of a body. Here we give the statement of Archimede's principle.

Statement:

The principle states that when a body immersed completely or partially in a liquid will experience an upthrust (upward force) equal to the weight of the liquid (or fluid) displaced by the body.

We can verify Archimedes principle from following examples:

1. Holding a test tube vertically place its closed end on the surface of water and push it down vertically in the water. You would feel as if some force is acting upward on the test tube. The deeper you push it, the greater will be the upward force acting on it. This upward force of water acting on the test tube is called upthrust.

2. A piece of cork having a density less than that of water needs to be held in water, if it is to remain in it. When released the cork will rise to the surface and will float. It is the upthrust (upward force) causing it to rise.

These quantities are listed in table 2.1 with the symbols.

Table 2.1: SI Base Units

No.	Quantity	Name	Symbol
1.	Time	second	s
2.	Length	metre	m
3.	Mass	kilogram	kg
4.	Amount of substance	mole	mol
5.	Thermodynamic temperature	kelvin	K
6.	Electric current	ampere	A
7.	Luminous intensity	candela	cd

Some measured length, time and mass are listed below:

Some measured lengths (approximate values)

Length	Metres
Farthest observed quasar	2×10^{26}
Andromeda galaxy	2×10^{22}
Radius of our galaxy	6×10^{19}
The nearest star (Proxima Centauri)	4×10^{16}
The orbit radius of planet Pluto	6×10^{12}
Radius of the sun	7×10^{8}
Radius of the earth	6×10^{6}
Radius of hydrogen atom	5×10^{-11}
Effective radius of a proton	2×10^{-15}

Some measured time intervals (approximate values)

Time interval	Seconds
Life time of proton	$> 10^{40}$
Age of universe	5×10^{17}
Time of earth's orbit around the sun (1 year)	3×10^{7}
Time of earth's rotation about its axis (1 day)	9×10^{4}
Time between normal heart beats	8×10^{-1}
Period of oscillation of 3 cm microwave	1×10^{-10}
Life time of least stable particles	$< 10^{-23}$

Some measured masses (approximate values)

Object	Kilograms
Known universe (estimate)	10^{53}
Our galaxy	2×10^{43}
Sun	2×10^{30}
Earth	6×10^{24}
Moon	7×10^{22}
Speck of dust	7×10^{-10}
Virus	1×10^{-15}
Uranium atom	4×10^{-26}
Proton	2×10^{-27}
Electron	9×10^{-31}

2.2 DERIVED UNITS

All physical quantities derived from the fundamental quantities are known as DERIVED QUANTITIES and their respective units are called DERIVED UNITS. Derived units are obtained by multiplication, of fundamental units. For example, volume of a box is equal to length × breadth × height, that is, meter × meter × meter = m^3. Meter is a fundamental quantity, so meter cube is a derived unit. Similarly, speed, acceleration, force, density, area etc. are derived quantities and their units are derived units. Some derived quantities and their derived units with their symbols are given below:

Some derived quantities

Quantity	Symbol	Unit	Unit Symbol
Speed	v	meter/second	m/s
Acceleration	a	meter/second^2	m/s^2
Volume	V	cubic meter	m^3
Force	F	newton	N(=kg m/s^2)
Pressure	P	pascal	Pa(=N/m^2)
Work	W	joule	J(=N.m)
Charge	q	coulomb	C(=A.s)

The advantage of SI System of units is that it makes calculations easier because smaller and bigger units can be obtained just by a simple division or multiplication by a factor of ten. For larger and smaller quantities we can use prefixes with the units. Prefixes for factors greater than unity have Greek roots,

and those for factors less than unity have Latin roots (except femto and atto, which have Danish roots). These prefixes are listed below:

SI Prefixes

Factor	Prefix	Symbol	Factor	Prefix	Symbol
10^{18}	exa	E	10^{-1}	deci	d
10^{15}	peta	P	10^{-2}	centi	c
10^{12}	tera	T	10^{-3}	milli	m
10^{9}	giga	G	10^{-6}	micro	µ
10^{6}	mega	M	10^{-9}	nano	n
10^{3}	kilo	k	10^{-12}	pico	p
10^{2}	hecto	h	10^{-15}	femto	f
10^{1}	deka	da	10^{-18}	atto	a

2.3 SIGNIFICANT FIGURES

The measured value of a physical quantity is always a number usually a decimal number. Let this physical quantity be the length of a rod which is measured by a scale graduated in centimetres. We take a large number of readings on the given rod. The average value of the length calculated on basis of these readings will be more accurate. Let this value be 320.5 cm. In this number (320.5) the digit on the extreme right to the decimal point which is (5) is uncertain. It is just an estimated value. The reason of uncertainity is that scale can read correctly upto units placed of a number where this as the digit (5) lies on the first place of decimal point and hence uncertain. The remaining digits 3, 2, 0 are known with certainty.

The digits 3, 2, 0, 5 in the number are significant figures. Thus the significant figures in a number are the digits which are known with certainty thus the digit on the extreme right which is uncertain. The same mesurement when expressed in metres would be 3.205m. In this number the significant figures are four that is 3, 2, 0, 5 in which the digits 3, 2, 0 are known with certainty whereas the extreme right digit (5) is uncertain.

Sometimes the measured value of a physical quntity can be written in more than one way. But only one of them is the correct way of writing the measured value which contains correct number of significant figures. For example suppose that the measured of the length of a rod by a scale graduated in centimetres is 120 cm when expressed in metres is 1.20 m. Mathematically the numbers 1.20 and 1.2 are the same and so 1.20m = 1.2m. If we record our measurement as 1.2m, it is incorrect. This number has two significant figures 1, 2. The scale used can read upto $\frac{1}{100}$ of a metre, as clear from the measured

value 1.20 in which (0) lies on the hundreth place $(\frac{1}{100})$. The recorded value 1.2m shows that the scale could read upto $\frac{1}{10}$ of a metre which is wrong.

Similarly if we write our measured values as 1.200m, which is also equal to 1.20m (1.200m = 1.20m), the recorded value (1.200m) is incorrect because it shows that the scale used could read upto $\frac{1}{1000}$ of a metre which is wrong.

Hence the correct way of writing the above measured value is 1.20m. It is neither written as 1.2m nor 1.200m.

Take another example, let our scale be graduated in millimetres. The length of a rod is measured to be 326.4 mm on this scale. In this number there are four significant figures 3, 2, 6, 4. The digits 3, 2, 6 are known with certainty while the digit 4 is uncertain.

Rules for finding significant figures.

(i) All non zero digits are significant. For example the number 239 has three significant figures 2, 3, 9.

(ii) Zeros lying between non zero digits are significant, for example the number **2009** has four significant figures.

(iii) All the zeros which locate the decimal point in a number less than (one) are not significant. For example the number 0.00786 has only three significant figures that is 7, 8, 6.

(iv) The zeros which are located immediately to the right of the decimal point are significant. For example the number **78.000** has five significant figures.

(v) Zeros locating the decimal point in a number greater than 1 (one) are not necessarily significant. For example, the number **500** has only one significant figure. In such cases scientific notations is particularly convenient from the point of finding the significant figures. Suppose that a certain distance of 1500m is known to four significant figures. Writing the number as 1500m creates a problem, because it implies that only two significant figures are known. In contrast the scientific notations of the above number that is 1.500×10^3m has the advantage of indicating that the distance is known to four significant figures.

(vi) When two or more numbers are used in a calculations, the number of significant figures with answers is limited by the number of significant figures in the original data. For example a rectangular gardens with sides 9.8m and 17.1m has an area of (9.8m) x (17.1m). A calculator gives 167.58m^2

for this product. However one of the original lengths is known only to two significant figures so the final answer is limited to only two significant figures and should be rounded off to give $170m^2$. In general when numbers are multiplicated or divided, the final answer has a number of significant figures that equals the smallest number of significant figures in any one of the original factors.

The number of significant figures in the answer to an addition or subtraction is also limited by the original data. Consider the total distance covered by a biker's trail that consists of three segments with distances of 2.5 km, 11 km and 5.26 km. The total distance covered is:

$$\begin{array}{ll} 2.5 & \text{km} \\ 11 & \text{km} \\ 5.26 & \text{km} \\ \hline 18.76 & \text{km} \end{array}$$

The distance of 11 km contains no significant figures to the right of the decimal point. Therefore the sum of the distances also should not have significant figures to the right of the decimal point. Thus the total distance should not be taken as 18.76 km. Instead the answer is 19 km after rounding off.

(vii) Addition or subtraction of numbers expressed in scientific notations requires that they should be written with the same power of 10. For example

$$\begin{aligned} 2.25 \times 10^6 + 6.4 \times 10^7 &= 2.25 \times 10^6 + 64 \times 10^6 \\ &= (2.25 + 64) \times 10^6 \\ &= 66.25 \times 10^6 \end{aligned}$$

In the original data the minimum number of significant figures is 2. Hence the answer of addition of the above numbers is 66×10^6 or 6.6×10^7.

(viii) Significant zeros:

We shall explain the meaning of significant zeros by taking an example. Consider a number 1200. This number may have two, three or four significant figures, depending upon whether the zeros represent measurements or are marely used to locate the decimal point. Scientific notation avoids this ambiguity. The above number can be written as 1.2×10^3, 1.20×10^3 and 1.200×10^3. These numbers written in scientific notation have two, three and four significant figures respectively.

2.1 Solved examples:

Radius of a sphere = 0.20 cm
Given π = 3.14

Find the surface area of the sphere and express the answer in significant figures.

Solution

The surface area of a sphere is given as,

$$A\text{ (Surface area)} = 4\pi r^2$$

where r stands for the radius of the sphere.

$$\text{Surface area } A = 4 \times 3.14\,(0.20)^2$$
$$= 0.502654824 \text{ cm}^2$$

Here we see that calculator displays nine digits on the right of decimal point, but the least significant figures in the given data are two only i.e. 2 and 0. Therefore answer should be given upto two significant figures. Thus surface area = 0.50 cm^2.

2.2 Solved Example

Find the volume of a rectangular box whose dimensions are given below:

Length of a box = 2.00 cm
Breadth of the box = 1.5 cm
Height of the box = 1 cm

Solution

The volume of the box is given as,

$$V = l \times b \times h$$

Putting the values in the above formula

$$V = 2.00 \times 1.5 \times 1 = 3.00 \text{ cm}^3$$

Here we see that the answer contains three significant figures. But in the data the quantity having least significant figures has only one significant figure. Thus the answer should be given in only one significant figure i.e. 3 cm^3.

2.3 Solved Example

A steel rod has a diametre of 2.52 cm. Express its diametre in (i) millimetre (ii) and metre.

Solution

(i) Since

$$1 \text{ cm} = 10 \text{ mm}$$
$$\therefore\ 2.52 \text{ cm} = 25.2 \text{ mm}$$

It has three significant figures

(ii) Again

$$1\ \text{cm} = 0.01\ \text{m}$$
$$\therefore\ 2.52\ \text{cm} = 0.0252\ \text{m}$$

It has also three significant figures

2.4 Solved Example

In an experiment the mass of a calorimetre has been found 36.35g. Express this mass in (i) milligram, (ii) microgram; and (iii) kilogram.

Solution

(i) Since

$$1\ \text{g} = 10^3\ \text{mg}$$
$$\therefore\ 36.35\ \text{g} = 36.35 \times 10^3 \times \text{mg}$$
$$= 36350\ \text{mg} = 3.635 \times 10^4\ \text{mg}$$

(ii) since

$$1\ \text{g} = 10^6\ \mu\text{g}$$
$$\therefore\ 36.35\ \text{g} = 36.35 \times 10^6 \times \mu\text{g}$$
$$= 36350000\ \mu\text{g} = 3.635 \times 10^7\ \mu\text{g}$$

It has four significant figures

(iii) since

$$1\ \text{g} = 10^{-3}\ \text{kg}$$
$$\therefore\ 36.35\ \text{g} = 36.35 \times 10^{-3} \times \text{kg}$$
$$= 0.03635\ \text{kg}$$

It has four significant figures

2.4 MEASURING INSTRUMENTS

Whenever we prepare to perform an experiment in a laboratory first of all we choose a suitable instrument to measure the physical quantity. This physical quantity may be a length, a mass or a duration of an interval. For the measurement of length we use metre stick and preferably other precision instruments like vernier callipers and micrometer screw gauge, to find mass a physical balance and a stop-watch to find the time interval.

Now we describe few of such instruments in some detail.

1. Vernier Callipers

A metre stick is graduated in millimetres, hence it can measure a distance upto 1mm. To measure distances smaller than this other instruments are used. Vernier Callipers is one of such instruments that can be used to measure a distance upto 0.05mm.

A vernier callipers consists of a rectangular steel bar whose one side is graduated in millimetres. This scale is known as main scale (MS). A small scale usually consisting of 10 divisions which slides over the main scale is known as vernier scale (VS) (20 division vernier scale is also in use).

The instrument has two jaws called callipers, shown in Figure 2.1, which enables it to measure the internal as well as the external diametre of a cylindrical object. A relatively thin flat rod is attached to the sliding scale on its back which enables it to measure the inner depth of the hollow cylinder.

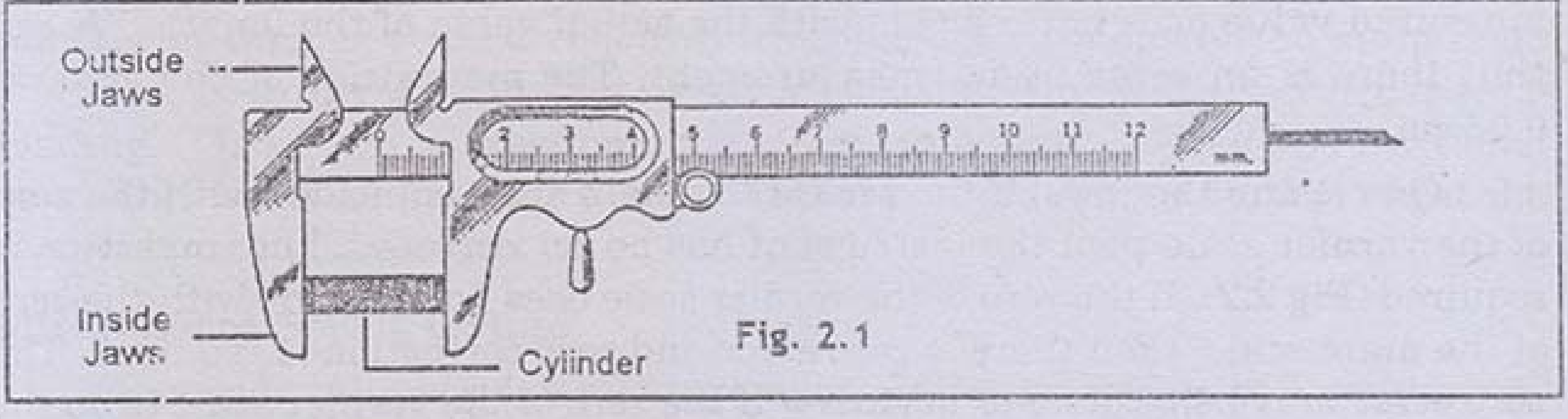


Fig. 2.1

Vernier Constant (VC) or Least Count (LC)

Vernier Constant (VC) or Least Count (LC) is the minimum distance that can be measured with the help of vernier callipers.

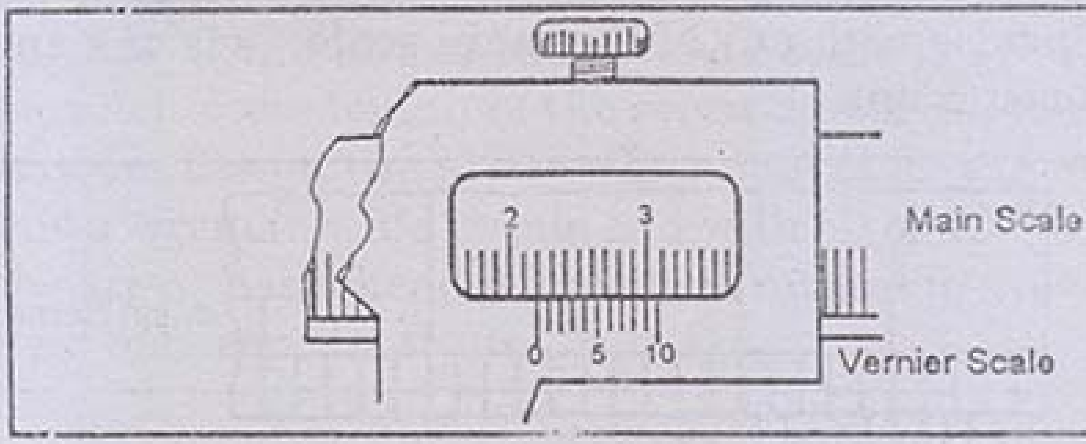


Fig. 2.2: Vernier Scale

It is obvious from the figure 2.2 that

10 vernier divisions = 9 main scale divisions

= 9 mm

$$1 \text{ vernier division} = \frac{9\text{ mm}}{10} = 0.9\text{ mm}$$

1 main scale division = 1mm

Least count = difference between 1 MS division and 1 VS division

$= 1\text{mm} - 0.9\text{mm} = 0.1\text{mm}$

$= 0.01\text{cm}$

The least count can also be calculated as follows:

$$LC = \frac{\text{Value of the smallest division on MS}}{\text{Total number of division on the VS}}$$

$$= \frac{1\text{ mm}}{10} = 0.1\text{mm} = 0.01\text{cm}.$$

Zero Error and its Correction

Consider a rod whose actual length is 3.25 cm. To measure the length of the rod experimentally we use a scale graduated in cm. We take a large number of readings. Suppose the average value of length comes to be 3.28 cm. The measured value does not coincide with the actual value of the length. We say that there is an error in our measurement. The magnitude of the error is 0.03 cm.

On closing the jaws, if the zero of the main scale coincides with the zero of the vernier scale then the instrument has no zero error and no correction is required (Fig 2.2). If the zero of the vernier scale does not coincide with the zero of the main scale, then there is zero error and zero correction is required. The zero error may be positive or negative. If the zero of the vernier scale is on the right of the zero of the main scale as shown in Fig 2.3, the zero error will be positive, because the observed value measured is more than the actual value. So the difference is to be subtracted from the observed value. To calculate the positive zero error, look at the vernier scale and note the number of divisions of the vernier scale which coincides with any of the main scale divisions and multiply this number by the least count.

Suppose 5th division of the vernier scale coincides with one of the main scale divisions (Fig 2.3), then the zero error is +0.05cm and the zero correction is –0.05cm. In this case the zero correction is negative.

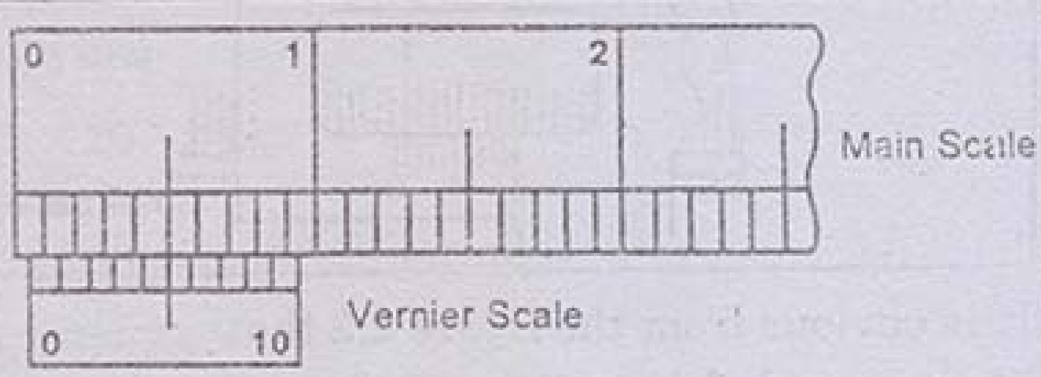


Fig. 2.3: Positive zero error

If the zero of the vernier scale is on the left of the main scale zero, then the zero error will be negative, because the observed measured value is less than the actual value. So the difference is to be added to the observed value. This zero correction is positive (Fig 2.4).

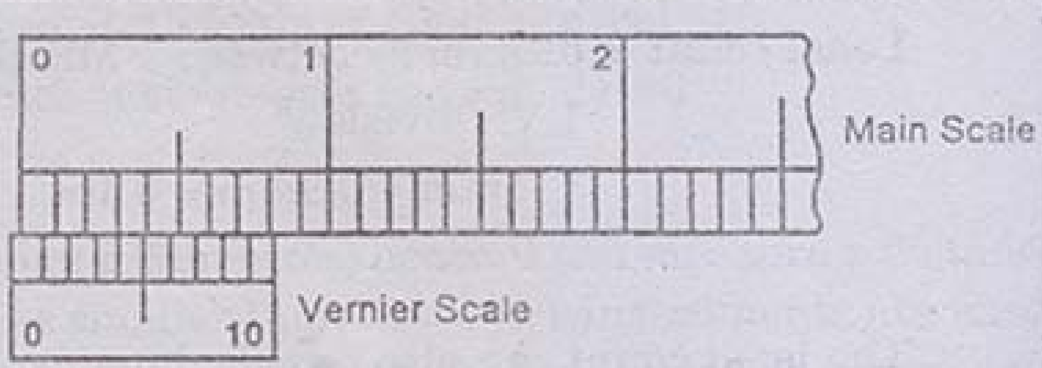


Fig. 2.4: Negative zero error

To find the negative zero error note the number of vernier scale division which coincides with any of the main scale division. If, say, it is 4, then the zero error will be $-(10-4) \times 0.01 = -0.06$cm. In this case zero correction is positive.

To measure the length of an object, it is placed between the two jaws of the vernier callipers as shown in Fig 2.1. The distance between the zero of the main scale and the zero of the vernier scale is equal to the length of the object. Note the main scale reading just before the zero of the vernier scale. Note the number of vernier scale division which coincides with one of the main scale divisions. Multiply this number by the least count. This will give vernier scale reading, it is known as fractional part. Add main scale reading and fractional part. This is the length of the object. If there is zero error, adjust it to get the correct length of the object.

Micrometre Screw Gauge

A vernier callipers can measure upto $\frac{1}{100\text{th}}$ or $\frac{1}{200\text{th}}$ of a centimetre. For more precise measurement micrometre screw gauge is used. A micrometre screw gauge consists of a U-shaped solid metal frame F. At one end of this frame a stud C with round end and flat face is fixed at A. A fine and accurately cut screw S having flat round end D passes through the other end B. On the outer surface of the nut of the screw a scale is graduated in millimetres. This is parallel to the length of the screw and is called the main scale which is linear. A drum fits on the screw which moves on the nut as it is rotated. This drum has a circular scale at one end with 50 or 100 divisions on it. At the other end the screw has a head R known as ratchet to avoid undue pressure on the object held between the studs C and D.

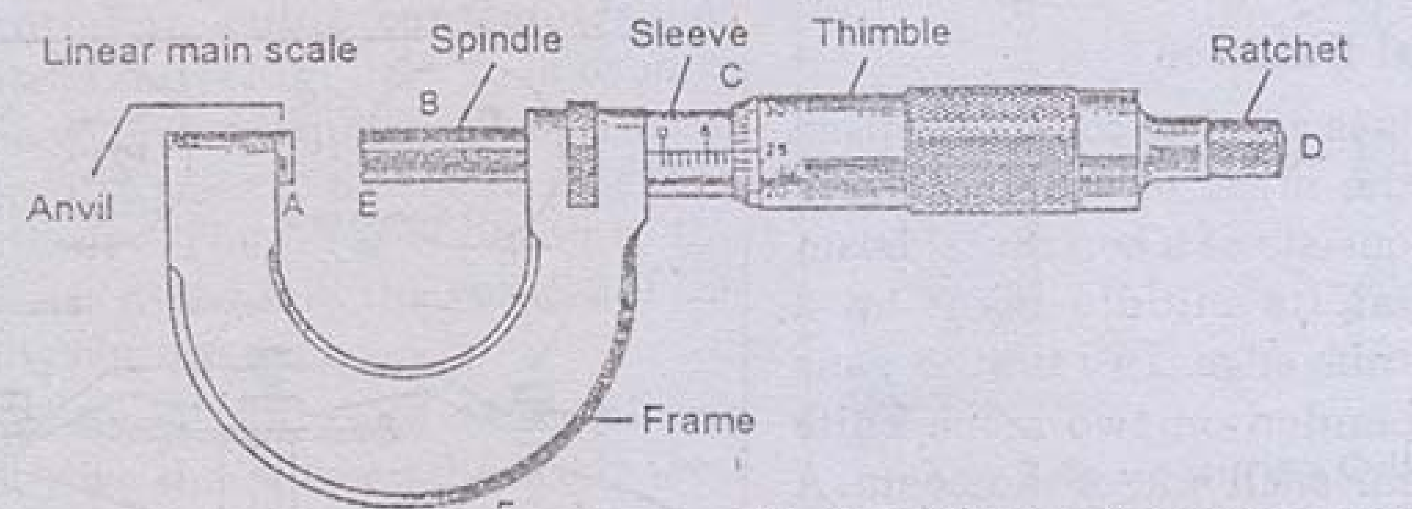


Fig. 2.5: Micrometer screw gauge

Pitch of the screw

It is the distance between the two consecutive threads of the linear screw. It is measured by the distance travelled by the circular scale on the main scale during one complete rotation of the circular scale. The least count of the screw gauge is given as,

$$\text{Least Count} = \frac{\text{Pitch of the screw}}{\text{Number of divisions on the circular scale}}$$

If the pitch of the screw is 1mm and the number of divisions on the circular scale is 100, then

$$LC = \frac{1\text{ mm}}{100} = 0.01\text{ mm}$$
$$= 0.001\text{ cm}$$

Zero Error and its Correction

Bring the studs C and D in contact by turning the screw. If the zero mark of the circular scale coincides with the reference horizontal line so that it coincides with the zero of the main scale, then there is no zero error. If the zero of the circular scale has advanced beyond the zero line of the main scale or left behind the zero of the main scale, there is zero error.

Bring the two studs in contact by turning the drum. If the zero of the circular scale is below the horizontal reference line, the zero error is positive, because it has measured more than the actual value. Note the number of divisions on the circular scale from zero which coincides with the reference line, multiply this number by the least count. This is positive zero error. Subtract this from the observed value to get the correct value. If the zero of the circular scale is above the reference line and the edge of the drum has crossed the zero line of the main scale, then the zero error is negative. Note the number of divisions on the circular scale from zero that coincides with the reference line, multiply it by the least count. This will be the negative zero error. Add this error to the observed value to get the correct measurement.

Physical Balance

This is a device commonly used to find the mass of an object (Fig 2.6). It consists of a horizontal beam resting at its middle point on a central knife edge. Two similar pans are suspended on two more knife edges near each end of the beam. A long pointer capable of swinging on a scale is attached to the middle of the beam. The physical balance is levelled on a table by means of levelling screws. The beam is set free by rotating the arresting knob at the front of the balance. The pointer is brought at the middle of the scale by means of two adjusting screws provided at each end of the beam.

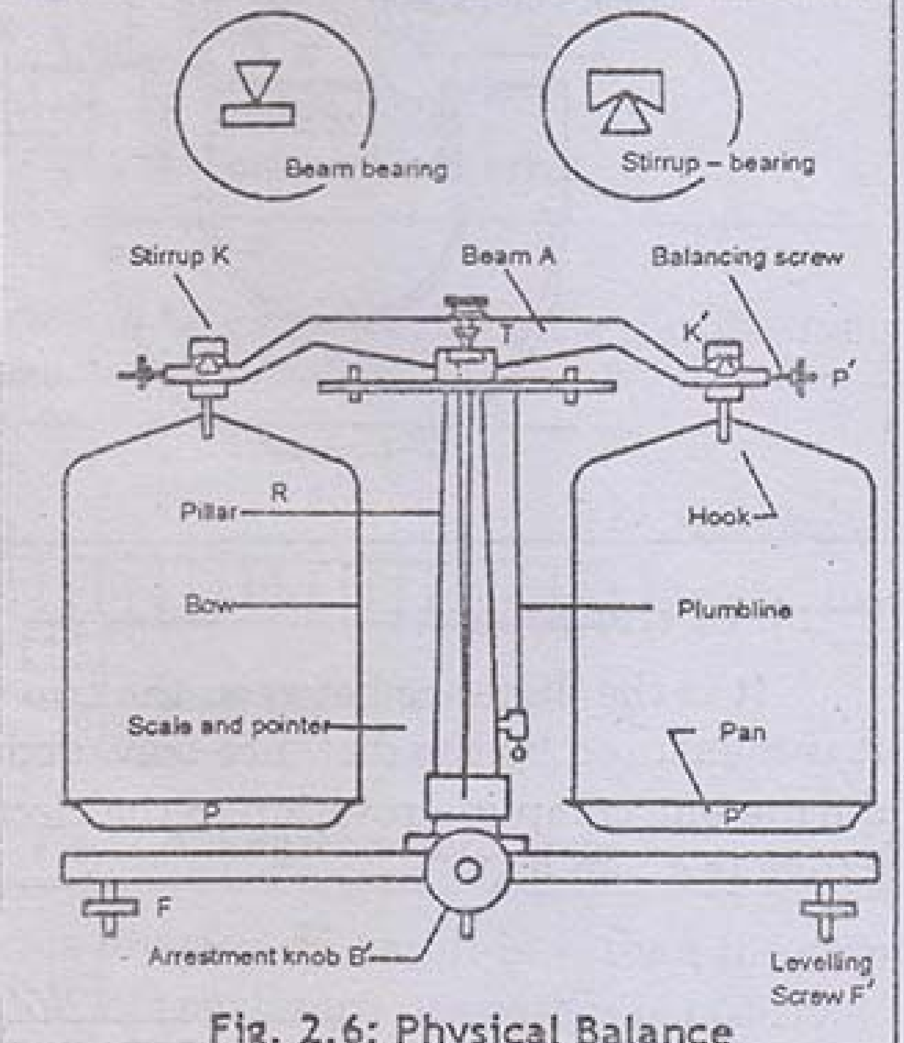


Fig. 2.6: Physical Balance

The arresting knob is turned to arrest the beam. The balance pans are cleaned. The object is placed on the left pan and the standard masses on the right pan. The beam is set free by turning the arresting knob. The pointer moves towards the side of smaller mass. The standard mass in the pan is adjusted to find the mass of the object.

Stop Watch

To measure time interval or to find instantaneous time we use a wrist watch or a clock and to keep its record we use a special watch known as a stop watch. Scales of minutes hand and seconds hand in it are on a circular dial (Fig 2.7).

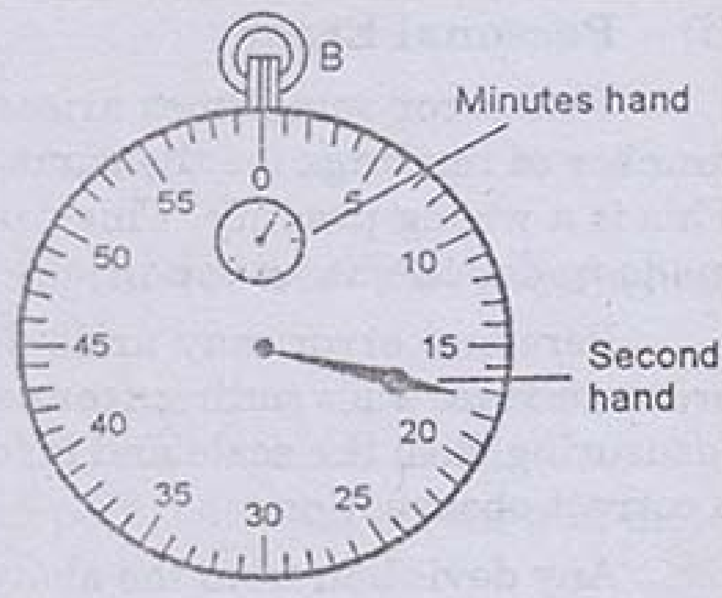


Fig. 2.7: Stop Watch

To note the time both hands of the stop watch are set at zero by pressing and releasing the knob B. As the knob B is pressed and again released the watch starts. When the second's hand completes one rotation of sixty seconds, the minutes' hand advances by one division. When we want to stop the watch, the knob B is pressed and released again. The new position of the hands gives the time interval for which the watch was in operation.

Measuring Cylinder

It is a glass cylinder with a scale-graduated in cubic centimetres or millilitres (ml) (Fig. 2.8). It is used to find the volume of liquids. When a liquid is poured, it rises to a certain height in the cylinder. The level of liquid in the cylinder is noted and the volume of the liquid is obtained.

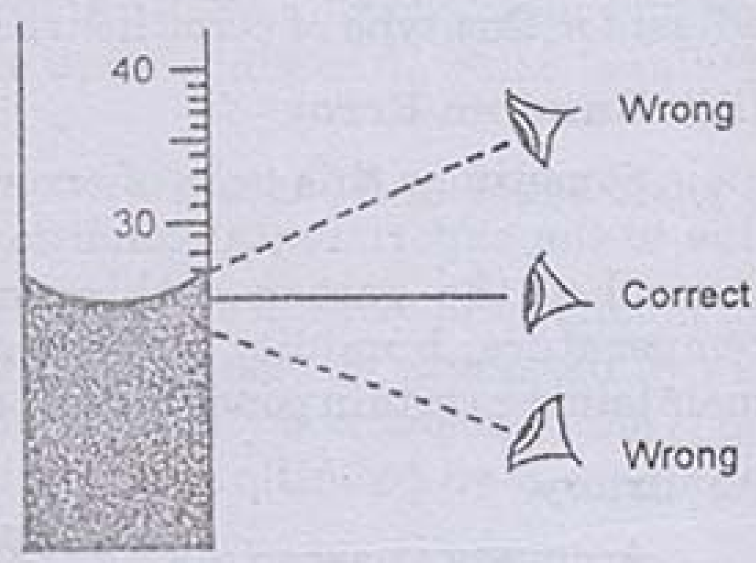


Fig. 2.8: Measuring Cylinder

In order to read the volume correctly we should keep the eye in level with the bottom of the meniscus of the liquid surface and the cylinder must be on a horizontal table.

Error and Accuracy

Whenever we perform experiments we use various instruments for the measurement of physical quantities. Such measurements have always certain error. An error is defined as the difference between the measured and the actual value. If two persons use the same instruments to measure the same physical quantity, they will not come up with the same results. Some observer with different instruments, the result of measurement will not be the same in general two considerations into account.

Classification of Errors

When we perform an experiment we come across with some errors. These errors can be put into three categories or types; (i) personal error, (ii) systematic error; and (iii) random error.

(i) Personal Error

This error sometimes arises due to the tendency of the person to make a number of readings for the same quantity in favour of one particular reading. This is a wrong practice. This practice should be avoided. All the observations made under the same conditions should be kept for further analysis.

Personal error may arise due to making an error in reading a scale. In order to record a reading from a scale we have to line up the object we are measuring with the scale and hold our eye in one particular position for making a correct observation.

Any deviation from the above procedure will result in an incorrect reading. The error discussed above come into play because of faulty procedure by the observer, so it is known as personal error.

(ii) Instrumental Error

This error is due to a fault in the measuring instrument as has already been discussed while describing the vernier callipers and micrometer screw guage. This is usually called zero error and may be positive or negative. To adjust for this type of error instructions have already been give above.

(iii) Random Error

Sometimes this type of error is also called accidental error. This arises due to the experimental conditions which are at a certain time beyond the control of the person making measurements. For example, changes in temperature, humidity, voltage etc. This type of error may be reduced by maintaining certain conditions in the laboratory.

Accuracy

Accuracy is ascertaining the measurement of the quantities involved in a phenomenon as close to the factual values as possible.

The experiments in laboratories can be done accurately if we can control the errors discussed above. Here it will be sufficient to say that one should try to reduce the errors and increase the accuracy of his experiments. If this can be done the results of the experiments that we perform will be more accurate and reliable. One way of doing this is to take a large number of readings of the same measurement and take their average as given below.

The average or mean value of the set of numbers 5.42, 6.18, 5.70, 6.01 and 6.32 is given as,

$$\text{Average or mean value} = \frac{5.42 + 6.18 + 5.70 + 6.01 + 6.32}{5}$$
$$= 5.93$$

It can be seen that 5.42 is the lowest number and 6.32 is the highest number and the deviation between these extreme numbers is 0.90. However, the deviations of the individual numbers from the mean or average value is less. Therefore, the average value is accepted close to the real value.

SUMMARY

1. **Base quantities:** Length, mass, time, temperature, electric current and light intensity.
2. **Derived quantities:** All physical quantities derived from the fundamental quantities are known as Derived Quantities.
3. **International system of units:** A system consisting of seven base units known as international system of units. It is abbreviated as SI.
4. **Base units:** Metre, kilogram, second, ampere, candela, kelvin and mole are base units of system international.
5. **Scientific notation:** An internationally accepted way of writing numbers in which numbers are recorded using the power of ten or prefixes and there is only one non zero digit before the decimal.
6. **Vernier calipers:** An instrument which can measure length correct upto 0.1 mm.
7. **Screw gauge:** This instrument can measure upto 0.01 mm.
8. **Uncertainty and error in measurement:** No measurement can be perfectly accurate. There remains some uncertainty in it, which is called error in measurement.
9. **Significant figures:** The accurately known digits and first doubtful digit in any measurement.
10. **Proportionality constant:** If two quantities are directly proportional to each other then their mutual relationship can also be expressed by an equation, while doing so the sign of proportionality is changed in the sign of equality and the independent variable is multiplied by a constant number known as constant of proportionality.

QUESTIONS

2.1 Write answers to the questions given below.

(i) What are fundamental and derived units?

(ii) With what instruments can you find

(a) mass of a body (b) length of a rod
(c) time of an experiment

(iii) Define

(a) standard of length (b) standard of time
(c) standard of current

(iv) Give a brief account for significant figures.

2.2 Fill in the blanks.

(i) In every day life __________ plays an important role.

(ii) Almighty Allah has created every thing in this world with the __________.

(iii) Physical units that are derived from fundamental units and called ________.

(iv) Kilogram is a unit of __________ in __________ system.

(v) If you wish to measure length with accuracy greater than 0.01cm you would use a __________.

(vi) The zero error of a measuring instrument can be __________ or __________.

2.3 Given below are a few possible answers to each statement; identify the correct one.

(i) One meter is equal to __________

(a) 10^4 mm (b) 10^3 mm
(c) 10^2 mm (d) 10^6 mm

(ii) 10^{-9} second is called a __________

(a) decisecend (b) millisecond
(c) microsecond (d) nanosecond

(iii) The fundamental unit of length in SI system is __________

(a) kilometer (b) metre
(c) yard (d) foot

(iv) The standard metre is made of __________ and is placed at the International Bureau of Weights and Measures in Sevece, near Paris.

(a) platinum and copper (b) iron and copper
(c) iron and iridium (d) platinum and iridium

2.4 Pick out true and false from the following.

(i) The fundamental unit of time in SI units is minute.

(ii) The SI unit of acceleration is meter per second per second.

(iii) 1 nanosecond = 10^{-8} second.

(iv) Hundredth of an mm can not be measured correctly with a micrometer screw gauge.

(v) A physical balance is a device used for measuring weight.

(vi) The fundamental unit of length in SI units is centimeter.

Problems:

2.1 (a) From diagram 1 calculate the zero error.

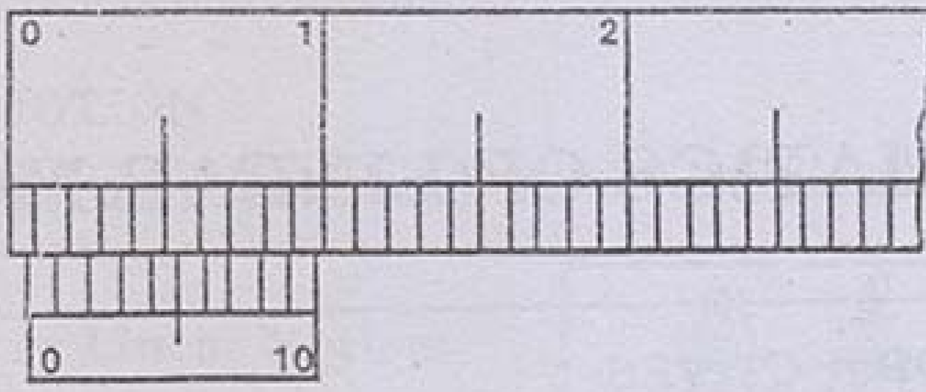


Diagram 1

(b) Diagram 2 shows the reading obtained for the diameter of a sphere using the same callipers. Calculate the diameter of the sphere.

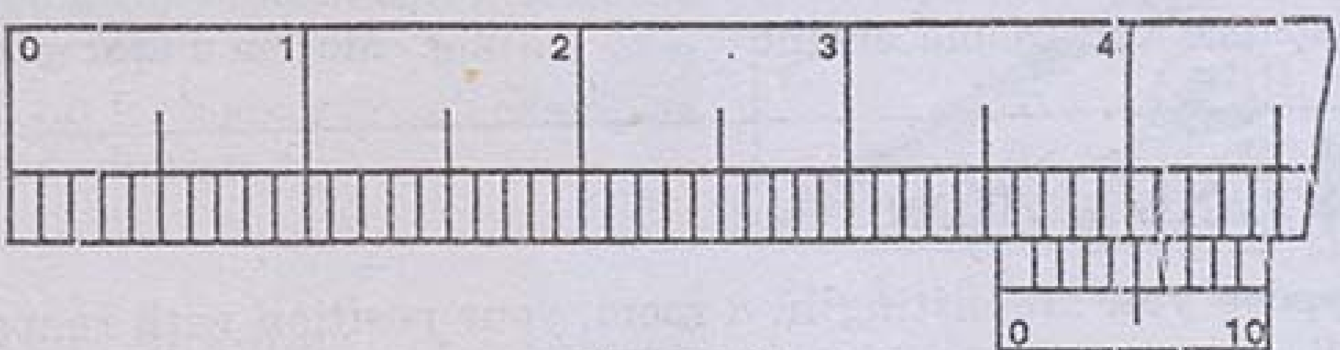


Diagram 2

2.2 The mass of an electron is 9.11×10^{-31} kg. Convert it in gm, milligram and microgram.

(Ans. 9.11×10^{-28} gm, 9.11×10^{-25} mg, 9.11×10^{-22} μgm)

2.3 The radius of hydrogen atom is 0.53×10^{-10} m. Convert it in cm, mm nm.

(Ans. 0.53×10^{-8} cm, 0.53×10^{-7} mm, 0.53×10^{-1} nm)

2.4 The time taken by an electron to complete one rotation about its nucleus is 0.5×10^{-18} seconds. Convert it into minutes, hours and microseconds.

(Ans. 8.333×10^{-21} minutes, 1.388×10^{-22} hours, 0.5×10^{-12} microseconds))

CHAPTER – 3

KINEMATICS OF LINEAR MOTION

LEARNING OBJECTIVES:

- Rest and motion.
- Types of motion.
- Kinematics.
- Scalar and vector quantities.
- Distance and Displacement.
- Speed
- Velocity
- Acceleration.
- Equations of motion.
- Motion under gravity.

3.1 REST AND MOTION

Suppose you are sitting in a room, your position with respect to the surrounding walls and the roof is not changing you are said to be in the state of rest. Similarly if you are standing at a bus stand near a bus in such a way that the position of the bus with respect to you is not changing, then the bus is said to be in the state of rest. Thus we conclude that

"A body is said to be in the state of rest, if it does not change its position with respect to its surroundings."

Suppose the bus starts moving toward or away from you in such a way that the position of the bus with respect to you changes continuously, then the bus is said to be in the state of motion.

"A body is said to be in the state of motion, if it changes its position with respect to its surroundings."

Sometimes it happens that a body is at rest with respect to some other body but at the same time it is in the state of motion with respect to another body.

For example if a person is sitting in a compartment of a moving train will find that all the things around him in the compartment are stationary, but for

a person standing on the platform all the things in the compartment are in motion.

From the above observations we can conclude that state of motion and rest are simply relative to one another.

3.2 TYPES OF MOTION

There are many types of motion, but the most common types are given below:

1. Translatory or Linear Motion:

Linear motion is also called the translatory motion. In this type of motion, a body moves in a straight or curved path. Every particle in the body is being displaced by the same amount. The motion of a car on a flat road, the motion of a motorbike on a circular road and the motion of a cricket ball hit for a sixer are all examples of translatory motion.

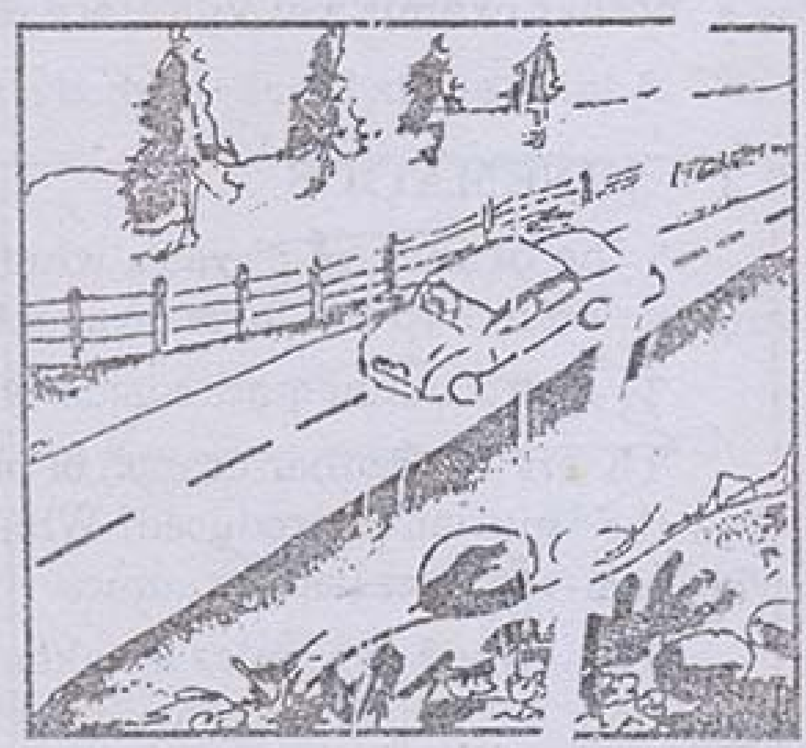


Fig. 3.1: Translatory Linear Motion

2. Rotatory Motion:

If a body spins or rotates about a fixed point or axis, its motion is called rotatory motion. The rotation of earth about its own axis, the wheels of a moving car and the blades of a moving electric fan are a few examples of rotational motion. The motion of every particle of the blades of the fan is a circular motion.

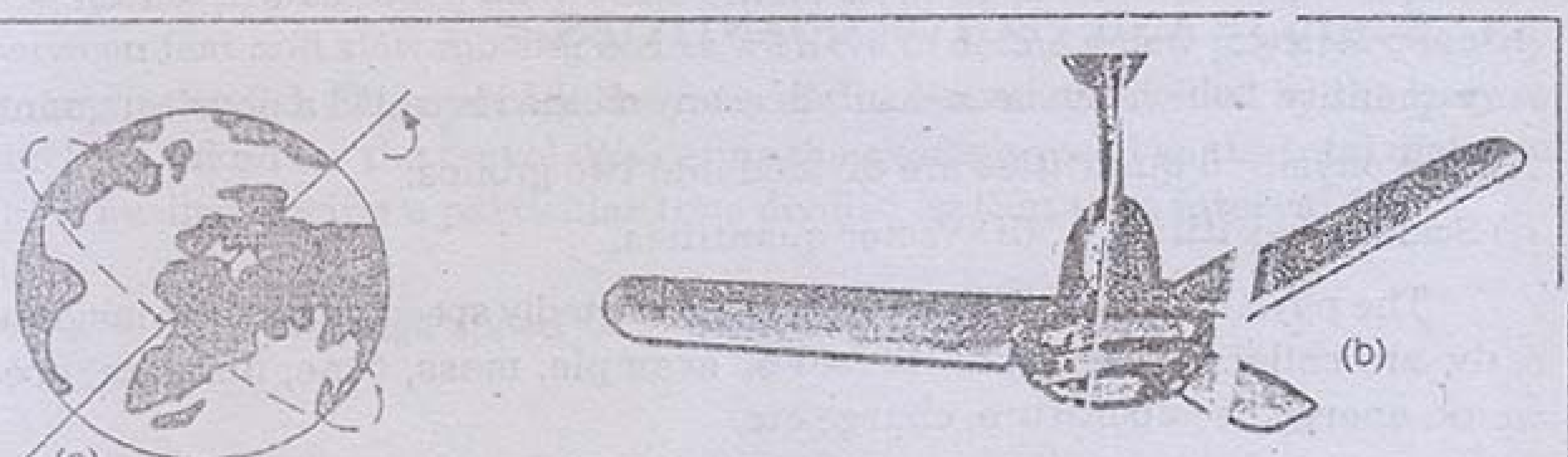


Fig. 3.2: Rotational Motion

3. Vibratory or Oscillatory Motion:

Consider a block which is suspended from a rigid support by means of a spring.

When the block is pulled in the downward direction (within the elastic limit) and released, it starts moving up and down about its mean position. This type of motion is known as a Vibratory or an oscillatory motion.

If a pendulum is desplaced from its rest position and released then it describes to and fro motion. This is another example of Vibratory motion.

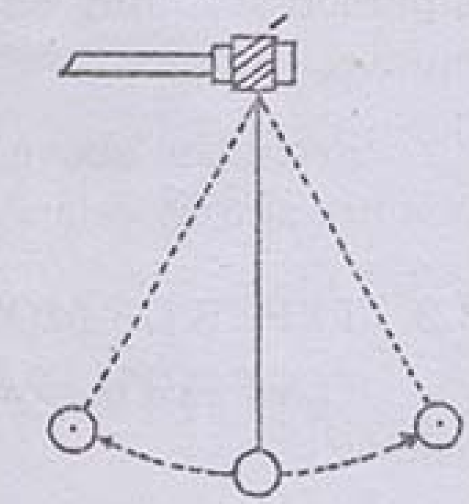


Fig. 3.3: Vibratory or Oscillatory Motion

3.3 KINEMATICS

The branch of physics which deals with the motion of objects with out any reference to the force or agent causing the motion is called kinematics.

The word kinematics is derived from Greek word *kinema* meaning motion.

There is another branch of physics which deals with the causes of motion. How the motion is produced? What are the factors which will affect the motion? This branch is called dynamics.

The word dynamics is taken from a Greek word *dynami* meaning power. So dynamics is the branch of physics which deals with the causes of motion and how they affect the motion.

In kinematics we study the position and motion of an object in space at a certain time without considering the causes of motion, while in dynamics we study the pushes or pulls which cause or resist the motion.

Mechanics is the branch of physics that deals with the kinematics and dynamics of objects.

3.4 SCALAR AND VECTOR QUANTITIES

Any quantity which can be measured by any method is called a physical quantity.

All the physical quantities are divided into two groups:

(i) Scalar quantities (ii) Vector quantities.

The physical quantities, which are completedly specified by their magnitude only, are called scalar quantities. For example, mass, time, distance, speed, work, energy, temperature, charge etc.

The scalar quantities can be added, subtracted and multiplied by simple arithmetic methods.

The physical quantities, which are completely specified by their magnitude and direction both, are known as vector quantities. For example, displacement, velocity, acceleration, force, momentum, torque etc.

The vector quantities acting along different directions can not be added, subtracted and multiplied by simple arithmatic methods. They are added, subtracted and multiplied by geometric or trigonometric methods.

A vector quantity can be represented graphically by a directional line segment. The length of the line segment gives magnitude whereas the arrow-head gives its direction.

3.5 DISTANCE AND DISPLACEMENT

Consider two points A and B such that a body is placed at the point A. The body can move from the point A to the point B along different paths I, II, III and IV, as shown in the fig. It can be seen from the fig that the body has travelled the distance in a straight line along the path IV only. The distance travelled by the body between the points A and B along the path IV is the shortest distance and is covered in a definite direction. The distance covered by a body in a particular direction is known as displacement.

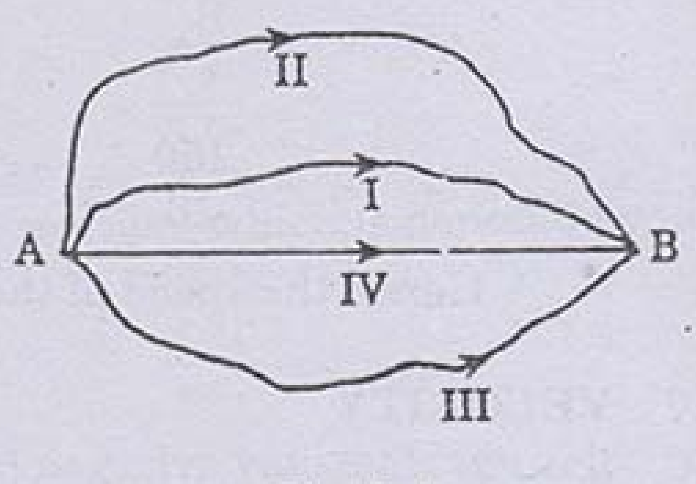


Fig. 3.4

Distance possesses magnitude only whereas displacement has magnitude as well as direction, hence distance is scalar and displacement is a vector quantity.

3.6 SPEED

Some bodies move fast, while others slow. In order to have a comparison between fast and slow moving bodies we have to define a new physical quantity known as speed. The speed is the ratio of the distance travelled by the body to the time taken for the travel. We define the average speed as the total distance "S" travelled during a particular time divided by that time interval "t".

$$\text{Average speed} = \frac{\text{Distance travelled}}{\text{Time taken}}$$

$$\text{i.e. } v_{av} = \frac{S}{t} \text{ ------ (1)}$$

Thus speed can be defined as distance covered by a body in unit time. In S.I units, the unit of speed is metre per second and is written as ms^{-1}.

When a body covers equal distance in equal intervals of time, however small the intervals may be, the speed of the body is said to be uniform.

Example:

A motor cyclist covers 150 m in 10 seconds. Find the speed of the motor cyclist.

Solution:

Distance covered by the motor cyclist = $S = 150$ m
Time taken by the motor cyclist = $t = 10$ seconds
Speed of the motor cyclist = $v = ?$

Formula:

$$\because \quad v = \frac{S}{t}$$

$$\therefore \quad v = \frac{150}{10} = 15\ ms^{-1}$$

Hence the speed of the motor cyclist is $15\ ms^{-1}$

3.7 VELOCITY

Speed refers only to the time rate at which a body is moving whereas velocity refers to both the speed and direction in which the body is moving. The velocity can be defined as the rate of change of displacement or the distance covered by the body in a definite direction.

If a body is displaced through a distance $\vec{S}$ during time t, its average velocity $\overrightarrow{v_{av}}$ is given by.

$$\overrightarrow{v_{av}} = \frac{\vec{S}}{t} \text{ ------- (2)}$$

When speed and direction of motion of a body do not change, the velocity is said to be uniform. The velocity of the body changes of either its speed changes or direction or both of them. Such a velocity is called a variable velocity.

For example, a car is moving on a straight path, sometimes it moves faster and sometimes slower, thereby its speed is changing. The velocity of the body is said to be changing though its direction remains same. If a body moves in a circle with a constant speed then magnitude of the velocity remains same but its direction continuously changes. In both these cases change in velocity takes place, that is the velocity is variable. Velocity is a vector quantity and its unit is meter per second written as m/s or ms^{-1}, in SI units.

3.8 ACCELERATION:

When the velocity of a body changes, on acceleration is always produced.

The time rate of change of velocity of a body is called its acceleration.

Consider a body whose velocity changes from $\vec{v_i}$ to $\vec{v_f}$ in a short interval of time Δ t. As velocity of the body is changing hence an acceleration $\vec{a}$ is produced in it.

Change in velocity of the body in time Δt = $\vec{v_f} - \vec{v_i} = \Delta \vec{v}$

$\therefore$ Change in velocity in unit time $= \dfrac{\Delta \vec{v}}{\Delta t}$

or Rate of change of velocity $= \dfrac{\Delta \vec{v}}{\Delta t}$

Since rate of change of velocity is called acceleration, which is denoted by $\vec{a}$.

$$\vec{a} = \frac{\Delta \vec{v}}{\Delta t}$$

Acceleration can have both positive and negative values. If the velocity of a body increases continuously then the acceleration is said to be positive or simply acceleration and its direction will be same as that of the motion. If the velocity of a body decreases continuously then the acceleration is said to be negative and is called deceleration or retardation and its direction is opposite to the direction of motion.

If the velocity of a body changes by equal amount in equal interval of time then its acceleration is said to be uniform and if a body moves with uniform velocity, then it has zero acceleration.

Acceleration is a vector quantity and its unit is m/s^2 or ms^{-2} in SI units.

Example:

The velocity of a motor car moving along a road increases from 10 ms^{-1} to 50 ms^{-1} in 8 seconds. Find its average acceleration.

Solution:

Initial velocity of the car = $v_i = 10\ ms^{-1}$
Final velocity of the car = $v_f = 50\ ms^{-1}$
Time taken by the car = $\Delta t = 8$ s
Average acceleration = $a = ?$
Change in velocity = $\Delta v = v_f - v_i$
$= 50\ ms^{-1} - 10\ ms^{-1}$
$= 40\ ms^{-1}$

$$\therefore \quad \text{Average Acceleration} = a_{av} = \frac{\Delta v}{\Delta t}$$

$$= \frac{40}{8}$$

$$= 5\ ms^{-2}$$

Example:

A car is moving with uniform acceleration and attains the velocity of 72 Kmh^{-1} in 5 min. Find acceleration of the car.

Solution:

Initial velocity of the car $= v_i = 0$ (car starting from rest)
Final velocity of the car $= v_f = 72\ Kmh^{-1}$

$$= \frac{72 \times 1000}{3600} \cdot \frac{m}{s}$$

$$= 20\ ms^{-1}$$

Time taken by the car $= \Delta t\ 5\ min = 5 \times 60s = 300s$
Acceleration of the car = a = ?

Formula:

$$a = \frac{v_f - v_i}{t} = \frac{20 - 0}{300} = 0.066\ ms^{-2}$$

Hence the acceleration of the car is $0.066\ ms^{-2}$.

3.9 EQUATIONS OF MOTION:

We shall now derive the simple equation of motion for bodies travelling along a straight line with uniform acceleration.

FIRST EQUATION OF MOTION:

Suppose a body is moving with initial velocity v_i and is under going uniform acceleration "a" for a time "t" such that its final velocity becomes v_f.

Change in velocity of the body in time $t = v_f - v_i$

Therefore change in velocity in unit time $= \frac{v_f - v_i}{t}$

As change in velocity in unit time (i.e. the rate of change velocity) is called acceleration.

Therefore $a = \dfrac{v_f - v_i}{t}$

or $v_f - v_i = at$

$$\boxed{\therefore \quad v_f = v_i + at} \text{ -- (1)}$$

This is the first equation of motion.

SECOND EQUATION OF MOTION:

Suppose a body starts with an initial velocity v_i and travels with uniform acceleration "a" for a period of time t. The distance covered by the body in this time is "S" and its final velocity becomes v_f.

Since the acceleration is uniform, the velocity of the body increases or decreases by equal amount in equal time intervals i.e. the velocity changes at a constant rate.

Therefore:

$$\text{Average velocity} = \frac{(v_i + v_f)}{2}$$

Distance Travelled = Average Velocity × time

$$\text{Therefore} \quad S = \frac{(v_i + v_f)}{2} \times t$$

From the first equation of motion, we have

$$v_f = v_i + at$$

$$S = \frac{[v_i + (v_i + at)]}{2} \times t$$

$$S = \frac{(v_i + v_i + at)}{2} \times t$$

$$S = \frac{(2v_i + at)}{2} t$$

$$S = \frac{2v_i t}{2} + \frac{at^2}{2} = \frac{\cancel{2} v_i t}{\cancel{2}} + \frac{at^2}{2}$$

$$\boxed{\therefore \quad S = v_i t + \frac{1}{2} at^2} \text{ -------------------------- (2)}$$

This is called second equation of motion.

THIRD EQUATION OF MOTION:

The third equation of motion can be obtained by using equation (1) and (2) and removing t from them.

From equation (1) we have

$v_f = v_i + at$

Squaring both sides of the above equation we have.

$$(v_f)^2 = (v_i + at)^2$$

or $$v_f^2 = v_i^2 + 2\,v_i\,(at) + (at)^2$$

or $$v_f^2 = v_i^2 + 2v_iat + a^2t^2$$

or $$v_f^2 = v_i^2 + 2v_iat + \frac{2}{2}a^2t^2$$

or $$v_f^2 = v_i^2 + 2a\,(v_it + \frac{1}{2}at^2)$$

From equation (2), we have

$$v_it + \frac{1}{2}at^2 = S$$

$$\therefore \quad \boxed{v_f^2 = v_i^2 + 2aS} \quad \text{.................................} \quad (3)$$

This is the third equation of motion.

Example:

A car is moving with a velocity of 36 Kmh^{-1} on a straight road. On the application of the breakes it comes to rest after covering a distance of 10m. Calculate the deceleration of the car.

Solution:

Initial velocity of the car = v_i = 36 Kmh^{-1} = $\frac{36 \times 1000\ ms^{-1}}{3600}$

= 10 ms^{-1}

Final velocity of the car = $v_f = 0$

Distance covered by the car = S = 10 m

Deceleration of the car = a

Using equation:

$$2as = v_f^2 - v_i^2$$
$$2 \times a \times 10 = (0)^2 - (10)^2$$
$$20a = -100 \quad = \quad a = -\frac{100}{20}$$
$$a = -5 \text{ ms}^{-2}$$

Hence the car is moving with a deceleration of 5 ms^{-2} and the negative sign indicates that acceleration is negative.

Example:

A bus is moving with a velocity of 60 Kmh^{-1}. When breakes are applied it comes to rest after two seconds. Find the distance travelled by it, before coming to rest.

Solution:

Data:

Initial velocity of the bus = v_i = 60 Kmh^{-1}

$$= \frac{(60 \times 1000)\text{m}}{3600 \text{ sec}}$$
$$= 16.66 \text{ ms}^{-1}$$

Final velocity of the bus = $v_f = 0$

Time = t = 2 seconds

Distance covered = S = ?

From the first equation of motion, we have

$$v_f = v_i + at$$
$$\therefore \quad 0 = 16.66 + a \times 2$$
$$-2a = 16.66$$
$$a = -8.33 \text{ ms}^{-2}$$

Minus sign indicates that the velocity is decreasing.

Now from the third equation of motion, we get

$$2as = v_f^2 - v_i^2$$
$$2 \times (-8.33) \times S = (0)^2 - (16.66)^2$$
$$-16.66 \text{ S} = -(16.66)^2$$
$$\text{or} \quad S = +\frac{(16.66)^2}{+16.66}$$
$$\therefore \quad S = 16.66 \text{ m.}$$

Hence the bus will cover a distance of 16.6 m before coming to rest.

3.10 MOTION UNDER GRAVITY

If an object is thrown vertically upward it rises to a particular height and then falls back to the ground. This is due to the attraction of the earth on the object which pulls the object towards the ground. During the upward motion of the object, this attraction causes deceleration in the object, whereas the object is accelerated during the downward motion. The attraction of the earth on the thrown object is called gravitational attraction or gravity and the change in velocity of the object, as a result of this attraction, is known as acceleration due to gravity.

It was found by Galileo Galilee that all bodies irrespective of their weights are attracted by the same acceleration.

To prove this he released iron balls of different masses from the learning tower of Pisa and found that all of them struck the ground at the same time.

The acceleration due to gravity is represented by "g" and its value is $9.8ms^{-2}$ directed downward towards the centre of the earth but for simplicity we will take the value of g equal to $10ms^{-2}$.

The equations of motion for the bodies moving freely under gravity can be written as

$$v_f = v_i + gt$$

$$h = v_i t + \frac{1}{2} g t^2$$

$$2gh = v_f^2 - v_i^2$$

For an observer on the surface of earth upward direction is usually taken as positive. By this convention, g being downward the value of g negative i.e. $-9.8ms^{-2}$. For a body moving upward v is positive and for a body moving downward v is negative while g is negative for both.

Examples:

A ball is dropped from a tower. It reaches the ground in 10 seconds. Calculate the height of the tower and the velocity with which it hits the ground.

Data:

Initial velocity of the ball = $v_i = 0$
Acceleration due to gravity = $g = 10\ ms^{-2} = -10\ ms^{-2}$
Time taken by the ball = t = 10 seconds
Height of the tower = h = ?
Final velocity of the ball v_f = ?

Solution:

To find "h" we will use equation:

$$h = v_i t + \frac{1}{2} gt^2$$

$$= 0 \times t - \frac{1}{2} \times -10 \times (-10)^2$$

$\therefore$ $h = -500$ metres

The negative sign shows that the displacement is in the downward direction.

Final velocity of the ball.
Also from the equations of motion, we get

$v_f = v_i + gt$
$= 0 - 10 \times 10$
$= -100\ ms^{-1}$

Here the negative sign shows that v_f is in the downward direction.

Height of the tower = 500 m
Velocity with which the ball strikes the ground = 100 ms^{-1}

SUMMARY

1. **Mechanics:** The branch of physics, which deals with the motion of bodies.
2. **Kinematics:** It is the study of motion of bodies without any reference to mass or force.
3. **Dynamics:** It is the study of motion of a body under the action of a force.
4. **Types of motion:** Following are the different types of motion:
   a) Translatory motion
   b) Rotatory motion
   c) Vibratory motion
5. **Position:** The distance and direction of a body from a fixed point shows its position.
6. **Displacement:** The shortest distance between the initial and final positions of a body is called displacement.
7. **Scalar and Vector quantities:** Scalar is such a quantity which can be described by a number, with suitable unit, without the mention of direction. Whereas, vectors are those quantities which can be described by a number, with a suitable unit and with the mention of direction.

8. **Speed:** Distance covered by a body in a unit time is known as speed.

9. **Velocity:** Rate of displacement with respect to time is known as velocity.

10. **Acceleration:** Rate of change of velocity of a body is known as acceleration.

11. **Equations of motion:** Following are three equations of motion:

$$V_f = V_i + at$$

$$S = V_i t + \frac{1}{2} a t^2$$

$$2aS = V_f^2 - V_i^2$$

12. **Motion under gravitational acceleration:** Gravitational acceleration acts on bodies vertically downward, and its magnitude is 10 ms^{-2}.

PROBLEMS

3.1 Find the time taken by sunlight to reach the ground if the distance between the sun and the earth is 1.5×10^8 Km. Velocity of light is 3×10^8 ms^{-1}

(Ans. 8 min 20 seconds)

3.2 A person hears the echo of his own sound from a distant hill after 2 seconds. How far away is the person from the hill, if the speed of sound is 330 ms^{-1}

(Ans. 330 m)

3.3 A car moving with a velocity of 36 kmh^{-1} is brought to rest in 5 seconds; calculate its deceleration.

(Ans. 2ms^{-2})

3.4 Find the acceleration of a body whose velocity increases from 11ms^{-1} to 33 ms^{-1} in 10 seconds.

(Ans. 2.2 ms^{-2})

3.5 A body starting from rest acquires a velocity of 10ms^{-1} in 5 seconds. Calculate (a) the acceleration (b) the distance covered by the body in 5 seconds.

(Ans. 2ms^{-2}, 25 m)

3.6 A car starts from rest and after 20 seconds its velocity becomes 108 Kmh^{-1}. Find the acceleration of the car.

(Ans. 1.5 ms^{-2})

3.7 The velocity of a motor cycle increases by an acceleration of $2ms^{-2}$ and becomes $20ms^{-1}$ in 5 seconds. Find the initial velocity of the car.

(Ans. 10 ms^{-1})

3.8 A bus is moving with a velocity of 72 Kmh^{-1}. On the application of the breakes its stops after covering a distance of 500 m. Calculate the deceleration produced by the breakes.

(Ans. 0.4 ms^{-2})

3.9 A car starting from rest attains a velocity of 20 ms^{-1} in 5 seconds. Find the distance covered by the car.

(Ans. 50m)

3.10 A stone is dropped from the top of a tower takes 5 seconds to reach the ground. Calculate the height of the tower (take $g = 10ms^{-2}$)

(Ans. 125m)

3.11 A boy throws a ball with a velocity of 20 ms^{-1}. Find the time elapsed between the throwing and catching the ball?

(Ans. 4s)

3.12 A stone is thrown vertically upwards with a velocity of 20 ms^{-1}. Find the maximum height reached by the stone and the total time of flight.

(Ans. 20m, 4s)

3.13 A stone is dropped from a height of 40 m.

(a) How much time will it take to reach the ground?

(b) With what velocity will it strike the ground?

(Ans. 2.83s, 28.3 ms^{-1})

CHAPTER – 4

MOTION AND FORCE

LEARNING OBJECTIVES:

- Force.
- Newton's Laws of Motion.
- Mass and weight.
- Tension in a string.
- Momentum.
- Friction.

In the previous chapter, we have studied kinematics i.e. motion of objects without reference to the force which causes the motion.

Motion was simply described in terms of displacement, velocity, acceleration. In this chapter we shall deal with the causes of motion. The branch of mechanics which deals with the study of the causes of motion and the way they affect the motion is called dynamics.

4.1 FORCE:

We observe in every day life that force is needed to move or stop a body e.g. if we want to move a football we must kick it. Similarly if we want to move a heavy object (say a box) from one place to another we have to push or pull it. If we push a truck at rest, it will remain at rest, although the force is applied on it to produce motion. Similarly force can be used to stop a moving body but sometimes it is not possible. Thus force can be defined as an agent which produces or tends to produce the motion in a body or which stops or tends to stop the motion of a body.

Force can also distort or tends to distort the shape of a body to which it is applied

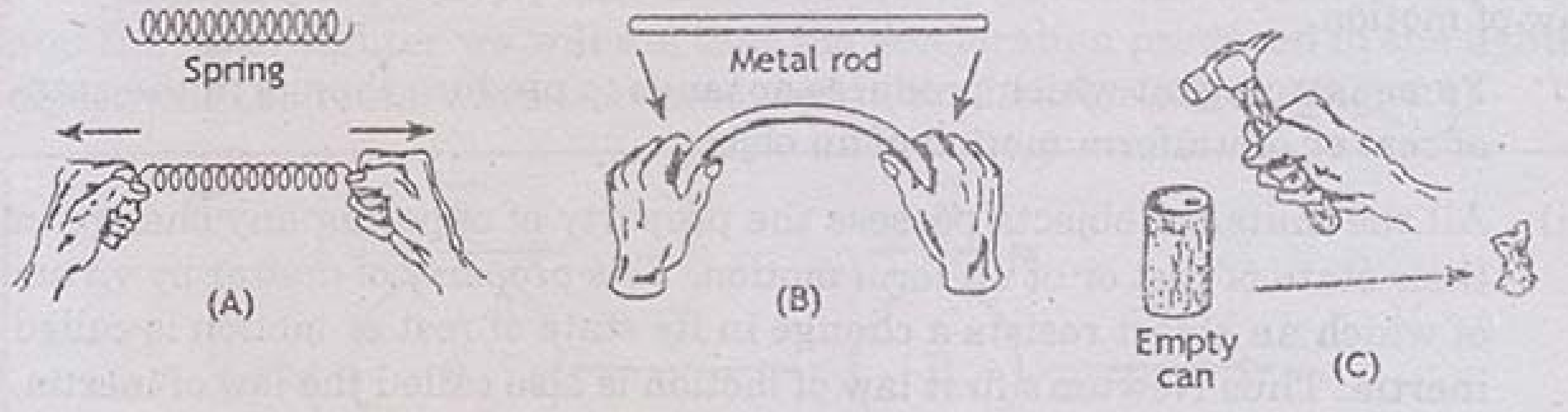


Fig. 4.1: Distorting effect of force

Above figures show that forces can increase or decrease the length of a spring, bend or twist a rod or change the shape of an empty can.

Thus force may also be defined as any agent which produces distortion in a body to which it is applied.

4.2 NEWTON'S LAWS OF MOTION:

The motion and its causes was one of the main problem for the physicsts. It remained unsolved for a long time. Galileo was amongst the earliest scientists who tried to solve this problem and pointed out that the motion of a body continues along a straight line unless there is some force that causes it to change the state of motion.

In 1687 Newton in his book "Philosophiae naturalis principia mathematica" meaning "The mathematical principles of natural philosophy" in which he mainly studied motion, of planets and other heavenly bodies. In his book Newton sums up the basic principles of motion in three laws which are as follows:

Newton's First Law of Motion:

We observe in every day life that an object at rest remains in the state of rest unless it is acted up on by some external force. For example, a book lying on a table can not change its position by itself unless a force is applied to change its position. Similarly if we push a ball on the ground it should continue its uniform motion indefinitely but it stops after covering certain distance. As soon as the ball starts moving, a force (force of friction) comes into play which opposes the motion of the ball.

These observations were summed up by Newton in his first law of motion which states that

"Every object continues its state of rest or of uniform motion in a straight line unless it is acted upon by an external force which changes its state of rest or of uniform motion".

The following two important conclusions can be drawn from Newton's first law of motion.

(i) Force is an agent which produces or tends to produce change in the state of rest or of uniform motion of an object.

(ii) All the material objects possess the property of opposing any change in their state of rest or of uniform motion. This property of matter by virtue of which an object resists a change in its state of rest or motion is called inertia. Thus Newton's first law of motion is also called the law of inertia.

Mass is the direct measure of inertia i.e. the greater the mass of an object the greater will be the inertia because relatively greater force is needed to change its state.

The following are few examples from our daily life which will help us in understanding inertia.

(1) A small coin is put on a card and placed over a glass. When the card is flicked away with the finger. The coin drops into the glass.

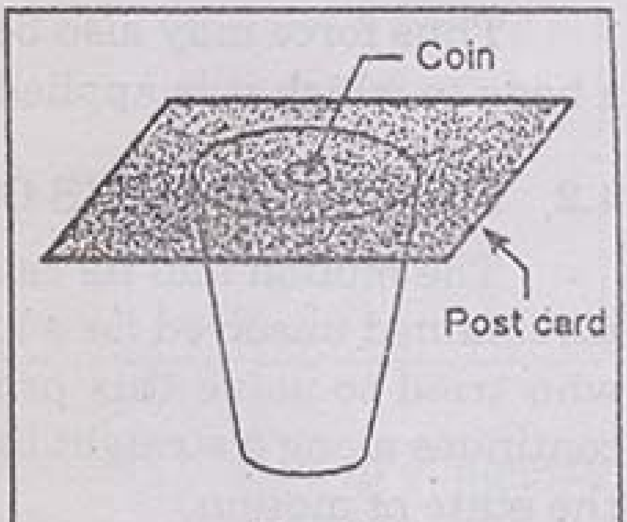


Fig. 4.2

(2) Suppose we are sitting in a stationary bus. If it starts moving suddenly we will feel a jerk in the backward direction. It is because our body is in contact with the seat of the bus and comes in motion with the motion of the bus while the upper portion of our body remains at rest due to inertia and so we feel a jerk in the backward direction.

Similarly if the bus is moving and stops suddenly we will feel a jerk in the forward direction. It is also due to inertia.

Newton's Second Law of Motion:

Newton's first law of motion has given us a simple definition of force as an agent that changes or tends to change the state of rest or motion of an object. For the determination of this force, through mathematical calculation, Newton derived a definite relationship between the force acting on an object, the mass of the object, and the acceleration produced. (Fig. 4.3)

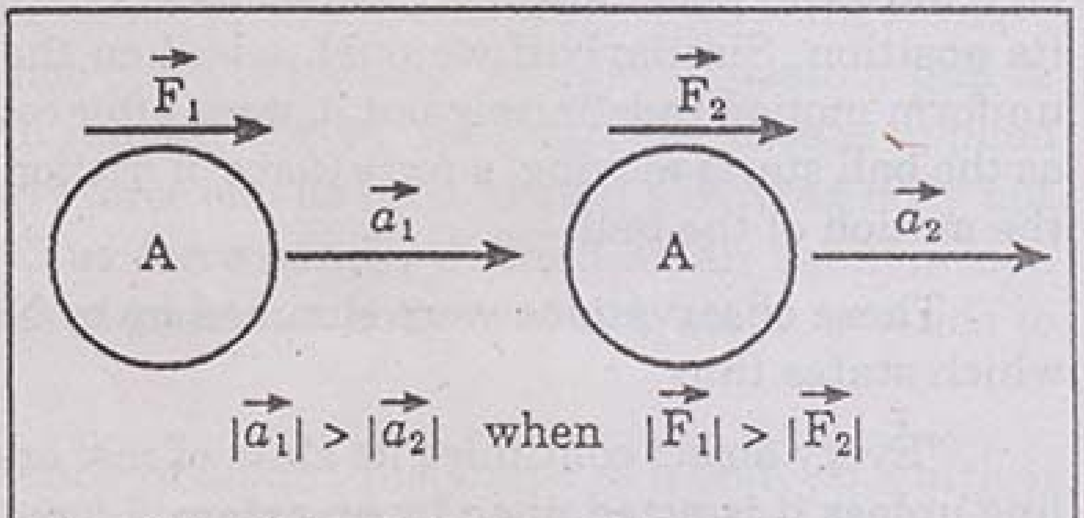


Fig. 4.3: Different force acting on the same mass

Similarly if we apply the same force on two different objects one heavier and the other lighter we will see that the acceleration produced in the lighter object will be greater then that produced in the heavier object. (Fig. 4.4)

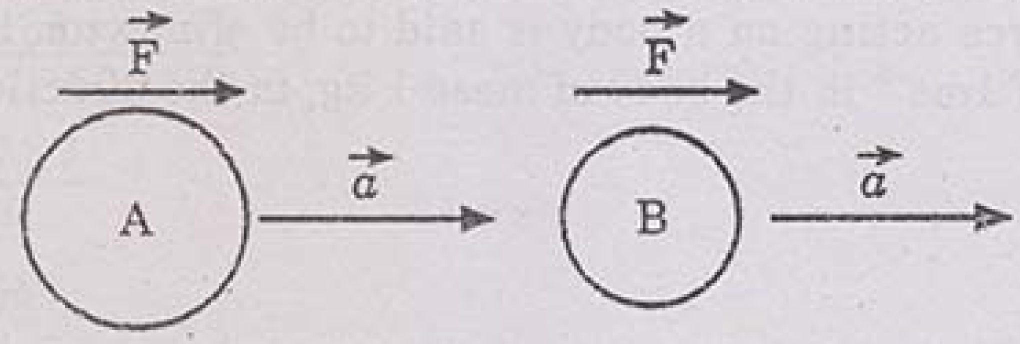


Fig. 4.4: Same force is acting on different mass

Newton summarised these results in the form of a law which is known as Newton's Second Law of Motion. It is stated as,

"When a force acts on an object, it produces acceleration in the object in its own direction. This acceleration is directly proportional to the magnitude of the applied force.

If a force "F" is applied on an object of mass "m" which produces acceleration "a" in the direction of the applied force then the relationship between acceleration and force can be mathematically expressed as:

$$a \propto F \text{ .... } (4.1) \quad \text{(for constant mass)}$$

Similarly for acceleration and mass the relationship can be written as:

$$a \propto \frac{1}{m} \text{ .... } (4.2) \quad \text{(for constant force)}$$

Combining equations (4.1) and (4.2) we get:

$$a \propto \frac{F}{m}$$

or $$a = K.\frac{F}{m}$$

Where K is constant of proportionally

or $$ma = KF \text{ .... } (4.3)$$

Irrespective of any system of units, if the acceleration produced in a body of unit mass is unity when a force of unit magnitude is applied, then obviously K = 1 and so

$$F = ma$$

In the System International, the unit of mass is kg and that of acceleration is ms^{-2}; the corresponding unit of force is kg m/s^2 which is termed as netwon (N).

"Thus force acting on a body is said to be of newton if it produces an acceleration of $1ms^{-2}$ in the body of mass 1 kg, in the direction of the applied force.

Example 4.1

A body of mass 50 kg is moving with an acceleration of 5 ms^{-2}. Find the force acting on the body.

Data:

Mass of the body = m = 50 kg
Acceleration produced in the body = a = 5 ms^{-2}
Force acting on the body = F = ?

Solution:

According to Newton's second law of motion

$$
\begin{aligned}
F &= ma \\
&= 50 \text{ kg} \times 5 \text{ ms}^{-2} \\
&= 250 \text{ kg} - \text{ms}^{-2} \\
&= 250 \text{ N}
\end{aligned}
$$

Hence a force of 250 N is acting on the object.

Example 4.2

A car of mass 1000kg traveling at 100 Kmh^{-1} is uniformly brought to rest over a distance of 40 m. Find (a) the average deceleration (b) the average braking force in newton.

Data:

Initial velocity of the car

$$= v_i = 100 \text{ kmh}^{-1} = \frac{100 \times 1000}{3600} \text{ ms}^{-1} = 27.78 \text{ ms}^{-1}$$

Final velocity of the car = $v_f = 0$
Distance covered = S = 40 m
(a) Acceleration = a = ?
(b) Braking force = f = ??

Using equation of motion, we have

$$v_f^2 - v_i^2 = 2as$$

$$(0)^2 - (27.78)^2 = 2 \times 40 \times a$$

$$a = -\frac{27.78 \times 27.78}{80}$$

$$= \frac{-771.7}{80}$$

$$= -9.646$$

$$a = -9.65\ ms^{-2}$$

$\therefore$ Average deceleration = $9.65\ ms^{-2}$

(b) Mass of the car = 1000 kg

Acceleration = $-9.65 ms^{-2}$

The force acting = f = ma

$= (1000)(-9.65)$

$= -9650$ N

The negative sign tells us that force is opposing the motion.

The average braking force = -9650 N

Newton's Third Law of Motion:

Before describing Newton's third law of motion let us consider the following examples,

(i) Place your hand on a table and press it hard in the downward direction, you will feel that the table is also exerting a force on your hand in the upward direction. The force applied by your hand on the table in the down ward direction is called action while the push of the table on your hand in the upward direction is called reaction. (Fig. 4.5)

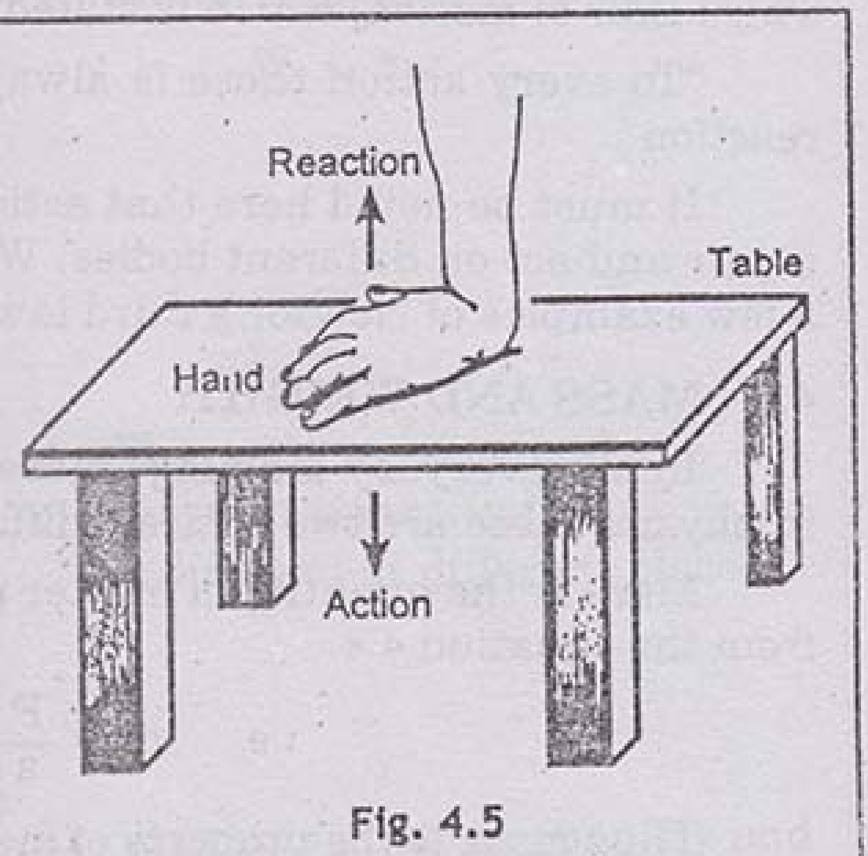


Fig. 4.5

(ii) Consider a book of weight W which is placed on a horizontal table It pushes the table with a force W in the downward direction. This is action of the book on the table. In response to this action the table also pushes the book in the upward direction. Since the book is at rest, the resultant force $\vec{F}$ acting on

the book must be zero that is

$$\vec{F} = 0 \Rightarrow F = 0$$

Two forces are acting on the book. they are:

(a) the force of gravity acting downward which we call the weight of the book.

(b) the other force is the push of the table on the book in the upward direction.

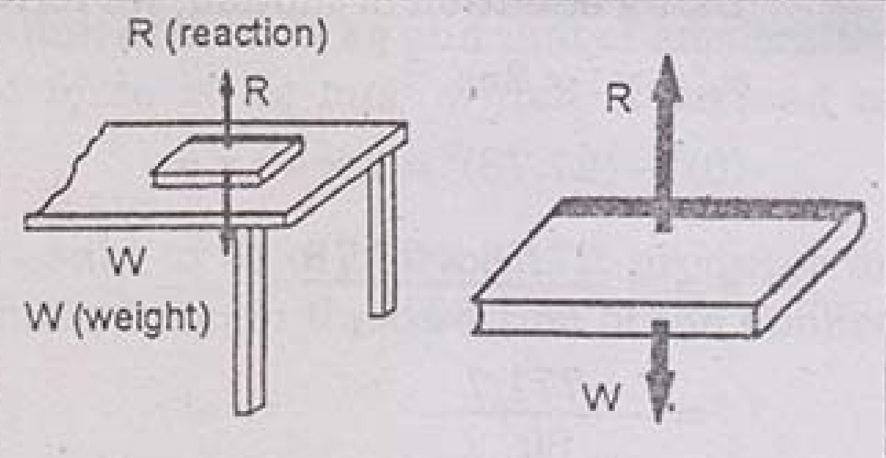


Fig. 4.6: Action and reaction forces between a book and a table

These two forces are equal and opposite to keep the book in the stationary condition.

In the engine of a rocket, gases formed due to the combustion of fuel, rush out with great speed through jet on the backside of the rocket and as a reaction the rocket moves in the upward direction.

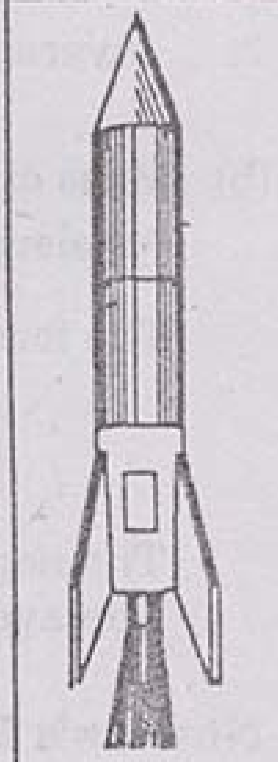


Fig. 4.7

From the above observtions it is clear that at least two forces are involved in the interaction between two bodies, one is called the action force and the other is called the reaction force.

This property of forces was summed up by Newton in his Third Law of motion, which is stated as,

"To every action there is always an equal and opposite reaction".

It must be noted here that action and reaction always occur together as a pair and act on different bodies. Walking, firing a bullet from a gun etc are a few examples of Newton's third law of motion.

4.3 MASS AND WEIGHT:

In our everyday language mass and weight carry the same meaning but in physics these are two entirely different quantities.

Mass is the quantity of matter possessed by a body and can be measured from the equation 4.4.

$$\text{i.e} \quad m = \frac{F}{a}$$

Thus mass is the property of matter which determines the acceleration of a body when acted upon by a given force. The greater the mass of the body the lesser will be the acceleration produced in the body on the application of a given force. Mass is the measure of inertia. It does not depend upon the location, shape, size of an object. It refers to the quantity of matter only. The more mass an object has, the more matter it contains. The unit of mass is kilogram and can be measured by physical balance.

All the bodies near the surface of the earth are attracted by the earth towards its centre with a certain force. This force of gravity acting on a body is called the weight of the object.

The direction of the weight of a body is always towards the centre of the earth. When a body is allowed to fall freely, the acceleration produced is called the acceleration due to gravity, which is denoted by 'g' and its weight W, is the force acting on it. According to Newton is second law of motion, we have

$$F = ma$$

For a freely falling body to

$$F = W \quad \text{and} \quad a = g$$

$$\therefore W = mg$$

Where "W" is expressed in newton, mass in kilogram and g in ms^{-2}.

COMPARISON BETWEEN WEIGHT AND MASS:

Weight and mass are two different quantities whose comparison is given in the following table.

Mass	Weight
1. The quantity of matter contained in a body is called its mass. It is the measure of inertia in a body. It is that property of a body which determines the acceleration produced in a body under the influence of given force force i.e. $a = \frac{F}{m}$	1. Weight is the force with which earth attracts a body towards its centre. Weight is given by $W = mg$
2. The mass of a body remains constant every where, whether it is measured at a point far away from the centre of the earth, or on the surface at the centre of the earth.	2. The weight of a body is not constant quantity but its value is different at different places.
3. Mass is a scalar quantity.	3. Weight is a vector quantity and is always directed towards the centre of the earth.
4. The S.I. unit of mass is kilogram.	4. The S.I. unit of weight is newton.

4.4 TENSION IN A STRING:

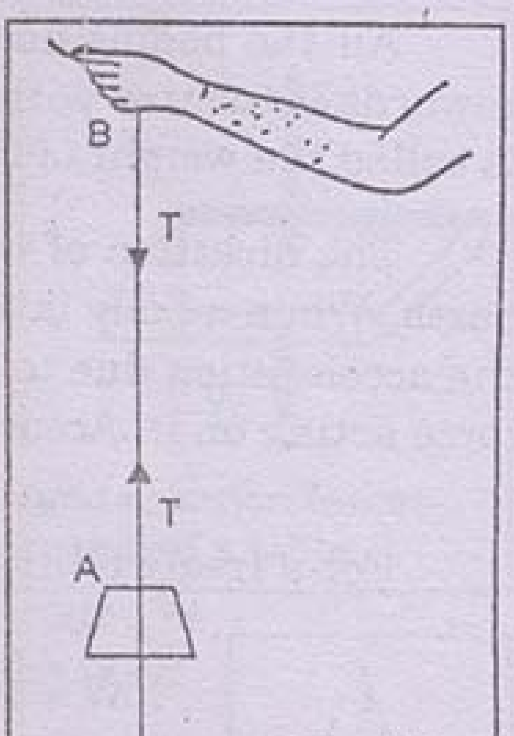


Fig. 4.8

Tension is a force exerted by a string on a body to which it is attached. When we fly a kite we feel some force enerted on our fingers. If we stretch a string tightly the hand will experience force acting towards each other. Similarly if a person is holding a block of weight W by means of a string, the weight of the block pulls it downwards, while the string pulls it upwards with an equal force. The force enerted on the fingers, the force experienced by the hands and the upward force acting on the block is due to the force developed in the strings in each case. This force is called the tension in the string.

In the Fig. 4.8 the hand experiences a pull in the downward direction at point B. Hence the tension at this point is downwards. However, at point A the string must exert a force upward to balance the weight of the body. Hence the direction of tension is upwards at point A. Thus the direction of tension changes but the magnitude of the tension remains constant at all points in the string. When the block is stationary, the magnitude of tension is equal to the weight of the block suspended from string. The S.I. unit of tension is newton.

MOTION OF BODIES CONNECTED BY A STRING PASSING OVER A FRICTIONLESS PULLEY:

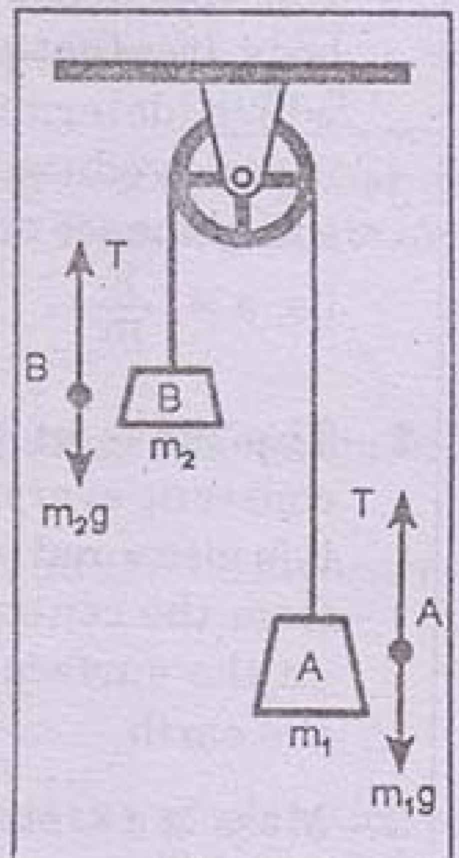


Fig. 4.9

Generally we come across two cases when we study the motion of bodies connected by a string passing over a pulley.

Case I: When both the bodies move vertically

Consider two bodies A and B having unequal masses m_1 and m_2 connected by a string which passes over a frictionless pulley in such a way that the two bodies hang vertically as shown in the Fig. 4.9.

Suppose m_1 is greater than m_2. Hence body A will move down with acceleration "a" while body B will move up with the same acceleration. Let T be the tension in the string.

To calculate the acceleration of the bodies and tension in the string, let us consider the motion of body A.

Two forces are acting on the body A.

(i) Weight of the body acting vertically downwards $= W_1 = m_1g$

(ii) The tension in the string acting vertically upwards $= T$

Since body A is moving down so W_1 is greater than T $(W_1 > T)$

The net force acting vertically downwards on body A $= F_1 = W_1 - T$

Or $= F_1 = m_1 g - T$ ..... (4.4)

Since the body A is moving down with acceleration 'a' the net force F_1 acting on A is given by Newton's second law of motion.

$$F_1 = m_1 a \quad \text{——————} \quad (4.5)$$

Comparing equations (4.4) and (4.5) we get

$$m_1 a = m_1 g - T \quad \text{——————} \quad (4.6)$$

Now consider the motion of body B, here also two forces are acting on it.

(i) Weight of the body W_2 acting vertically downwards. $= W_2 = m_2 g$.

(ii) Tension in the string T acting vertically upwards $= T$

Since the body B is moving vertically upward so tension in the string is greater than the weight of the body.

Thus net upward force F_2 acting on the body B is given by.

$$F_2 = T - m_2 g \quad \text{——————} \quad (4.7)$$

Applying Newton's second law of motion to body B.

$$F_2 = m_2 a \quad \text{——————} \quad (4.8)$$

Comparing equations (4.7) and (4.8) we get

$$m_2 a = T - m_2 g \quad \text{——————} \quad (4.9)$$

Adding equations (4.6) and (4.9) we get

$$m_1 a + m_2 a = m_1 g - T + T - m_2 g$$

or $m_1 a + m_2 a = m_1 g - m_2 g$

or $(m_1 + m_2)a = (m_1 - m_2)g$

$$\boxed{a = \frac{(m_1 - m_2)g}{(m_1 + m_2)}}$$

to calculate T divide equation (4.6) by (4.9) we get

$$\frac{m_1 a}{m_2 a} = \frac{m_1 g - T}{T - m_2 g}$$

Or $$\frac{m_1}{m_2} = \frac{m_1 g - T}{T - m_2 g}$$

$$m_1(T - m_2 g) = m_2 (m_1 g - T)$$

$$m_1 T - m_1 m_2 g = m_1 m_2 g - m_2 T$$

$$(m_1 T + m_2 T) = 2\, m_1 m_2 g$$

$$(m_1 + m_2)T = 2\, m_1 m_2 g$$

$$\boxed{T = \frac{(2m_1 m_2)g}{m_1 + m_2}}$$

Case II: When one of the bodies move vertically while the other horizontally

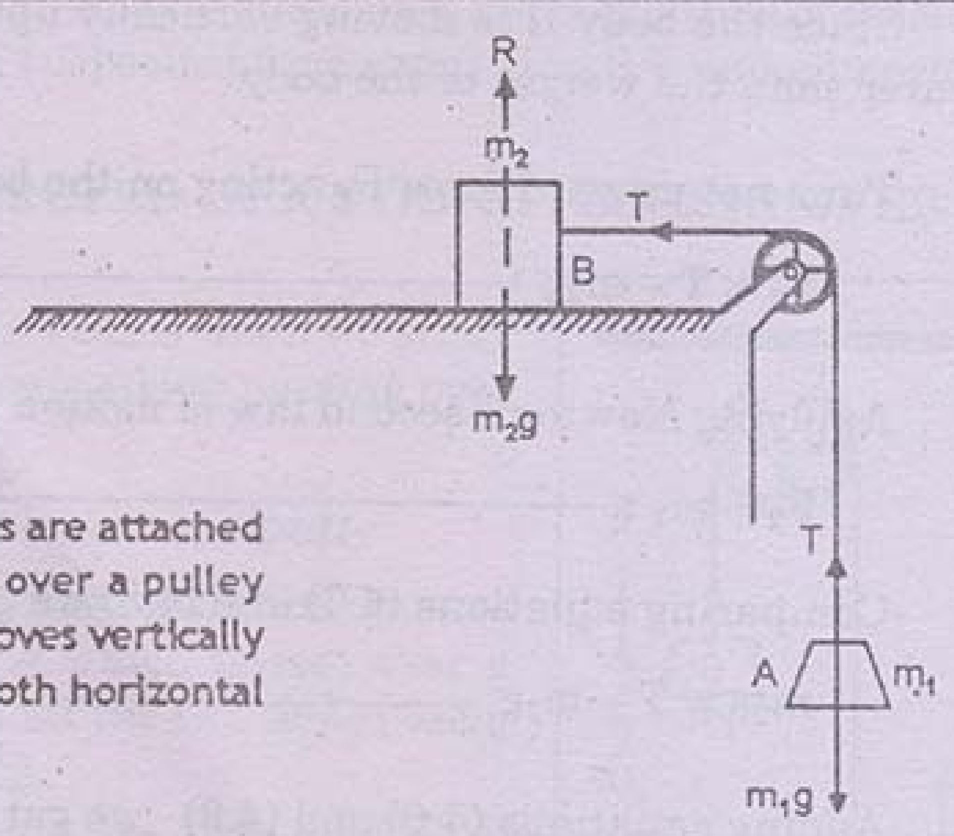


Fig. 4.10(a) Two unequal masses are attached to the ends of a string passing over a pulley in such a way that the body A moves vertically and the body B moves on a smooth horizontal plane.

Two bodies A and B of masses m_1 and m_2 respectively are connected to the ends of a string which passes over a pulley as shown in Fig. 4.10. The body A moves vertically downward with an acceleration "a" and the body B moves on a smooth horizontal table towards the pulley with the same acceleration.

Consider the motion of body A, two forces are acting on it.

(i) Weight "W_1" of the body A acting vertically downward.

(ii) Tension "T" in the string which is acting vertically upward. Since the body A is accelerated vertically downward so W_1 is greater than T.

So net downward force acting on the body A. = $F_1 = W_1 - T$

Or $\qquad F_1 = m_1 g - T$ —— (4.10)

On applying Newton's Second Law of Motion to body A we get

$$F_1 = m_1 a \quad \text{———— (4.11)}$$

Equating equation (4.10) and (4.11) we get

$$m_1 a = m_1 g - T \quad \text{———— (4.12)}$$

Now consider motion of the body B, three forces are acting on it.

They are:

(i) The tension "T" in the string acting horizontally towards the pulley.

(ii) The weight "W_2" acting vertically downward.

(iii) The normal reaction "R" of the smooth horizontal surface which acts vertically upward.

Since there is no motion of the body in the vertical direction, the two forces i.e. the weight W_2 and the reaction of the smooth surface R are equal and opposite. Hence they cancel each other.

Thus the only force acting on the body B is the tension T which is pulling the body towards the pulley.

$$F_2 = T \quad \text{———— (4.13)}$$

If the body is moving with acceleration "a" we get the value of force using Newton's Second Law of Motion.

$$F_2 = m_2 a \quad \text{(4.14)}$$

Comparing equations 4.13 and 4.14, we get

i.e. $$m_2 a = T \quad \text{———— (4.15)}$$

Adding equation (4.12) and (4.15), we get

$$m_1 a + m_2 a = m_1 g - T + T$$

$$m_1 a + m_2 a = m_1 g$$

$$(m_1 + m_2) a = m_1 g$$

Therefore $$\boxed{a = \frac{(m_1) g}{(m_1 + m_2)}}$$

To find the value of "T" put this value of "a" in equation (4.15) we get.

$$T = \frac{(m_1 m_2) g}{(m_1 + m_2)}$$

Example 4.3

Two bodies of masses 5 and 4 Kg are attached to the ends of a string which passes over a frictionless pulley such that the two bodies hang vertically. Find the acceleration of the bodies and tension in the string.

Let A and B be the two bodies so that

Mass of body A = m_1 = 5 kg

Mass of body B = m = 4 kg

$g = 10ms^{-2}$

Acceleration of the bodies = a = ?

Tension in the string = T = ?

Solution:

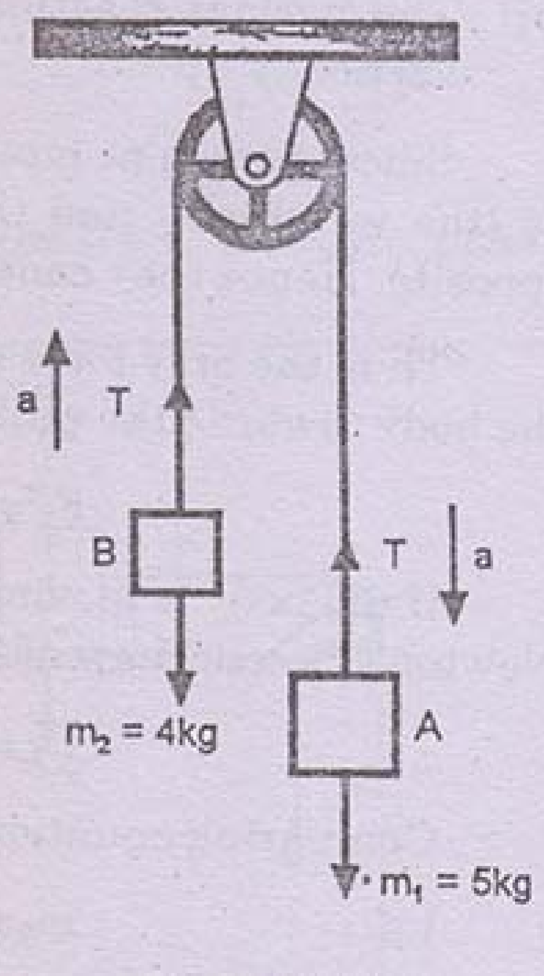


Fig. 4.10(b)

For calculation of "a"

$$a = \frac{(m_1 - m_2)g}{m_1 + m_2}$$

$$= \frac{(5-4)\,kg}{5+4} \times 10\ m/s^2$$

$$= \frac{1}{9} \times 10\ m/s^2$$

$$\boxed{a = 1.1\ ms^{-2}}$$

$$T = \frac{(2m_1 m_2)g}{(m_1 + m_2)}$$

$$= \frac{(2 \times 5 \times 4)}{5+4} \times 10$$

$$= \frac{400}{9} = 44.4\ N$$

Hence the acceleration of the bodies is $1.1\ ms^{-2}$ and the tension in the string is 44.4 N.

4.5 MOMENTUM:

It is more difficult to stop a loaded cart than the unloaded cart moving with the same velocity. Similarly we require a large force to accelerate a stationary loaded cart than a stationary unloaded one to the same velocity. The heavier body is said to possess greater quantity of motion than the body having lesser mass.

Similarly if we want to stop two identical balls moving with different velocities, we have to apply greater force to the ball moving with greater velocity than the ball moving with lesser velocity. It means that the ball moving with greater velocity possesses more quantity of motion than the ball heaving lesser velocity.

From the above observation we can conclude that the quantity of motion contained in a body depends upon mass of the body as well as the velocity of the body. This physical quantity is known as Momentum which is defined as the total quantity of motion contained in a body.

Mathematically momentum can be defined as the product of mass and velocity. If "m" is the mass of a body moving with velocity "v", then its momentum "p" given by :

$$\boxed{P = m\,v} \longrightarrow (4.16)$$

Momentum is a vector quantity. The S.I. unit of momentum is kg ms^{-2} (kilogram meter per second) or N-S (newton-Second)

LAW OF CONSERVATION OF MOMENTUM

Before studying law of conservation of momentum let us define an isolated system.

"When a number of bodies are such that they can exert force upon one another but no external agency exerts a force on them, they are said to constitute an isolated system of interacting bodies. For example, the molecules of a gas enclosed in a glass vessel at constant temperature. A perfectly isolated system is not possible.

The Law of Conservation of momentum states that the momentum of an isolated system always remains constant (or conserved). Consider a system of two non-rotating colliding balls A and B having masses m_1 and m_2 respectively. According to the conservation law of momentum

"The total momentum of the balls before collision = total momentum after collision".

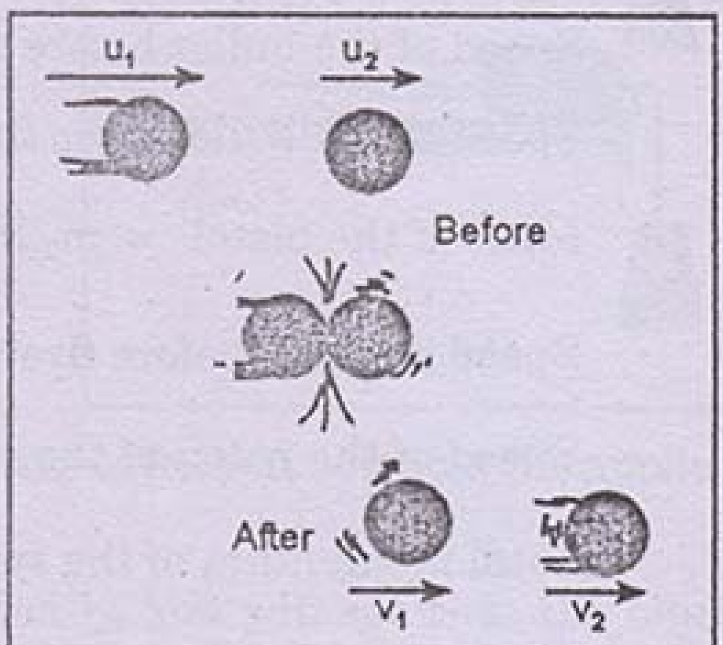


Fig. 4.11: Balls of masses m_1 and m_2 before collision then colliding with each other and finally after collision.

Let the balls A and B be moving along the line joining their centres in the same direction with velocities u_1 and u_2 respectively

such that u_1 is greater than u_2. Suppose the balls A and B acquire velocities V_1 and V_2 respectively after collision, as shown in Figure 4.11.

Momentum of the system before collision $= m_1u_1 + m_2u_2$

And momentum of the system after collision $= m_1v_1 + m_2v_2$

According to the law of conservation of momentum

Total momentum of the balls before collision = Total momentum after collision

Therefore $\boxed{m_1u_1 + m_2u_2 = m_1v_1 + m_2v_2} \longrightarrow 4.15$

APPLICATIONS OF LAW OF CONSERVATION OF MOMENTUM

If a balloon is blown up and released it flies round the room. During the flight, air escapes from the balloon in one direction while it moves in the opposite direction. This is due to the law of conservation of momentum of the system of the air and the balloon.

The recoil of a gun when a bullet is fired from it is due to the conservation of the momentum of the system, consisting of the gun and the bullet.

Example 4.4

A 5g bullet is fired from a 800g pistol with speed of 50 ms^{-1} What is the speed of the recoil of the pistol.

Solution:

Mass of the bullet $= m_1 = 2g = \frac{2}{1000}$ kg $= 0.002$ kg.

Speed of the bullet before fire $= u_1 = 0$

Speed of the bullet after fire $= V_1 = 50\ ms^{-1}$

Mass of the pistol $= m_2 = 800$ g $= \frac{800}{1000}$ kg $= 0.8$ kg

Speed of pistol before fire $= u_2 = 0$

Speed of the recoil of the pistol $= V_2 = ?$

Total momentum of the system before fire $= m_1u_1 + m_2u_2$
$= 0.002 \times 0 + 0.8 \times 0 = 0$

And total momentum of the system after fire $= m_1v_1 + m_2v_2$
$= 0.002 \times 50 + 0.8\ v_2$
$= 0.1 + 0.8\ v_2$

According to the law of conservation of momentum, we have total momentum after fire = total momentum before fire

$$0.1 + 0.8\,v_2 = 0$$

$$0.8\,v_2 = -0.1$$

$$v_2 = -\frac{0.1}{0.8} = -0.125\text{m/s}$$

$\therefore$ the speed of the recoil of the pistol = 0.125 m/s

4.6 FRICTION:

While dealing with motion we come across a term known as friction. The force which opposes the motion of a body while in continuous contact with the other body is called force of friction.

Friction plays a very important role in our daily life e.g. it allows us to walk and is necessary for the motion of wheels in the cars.

Friction is due to the roughness of two surfaces so it is unavoidable because there is no perfect smooth surface. When an object rubs against another, the roughness of their surfaces prevent them from sliding freely over one another.

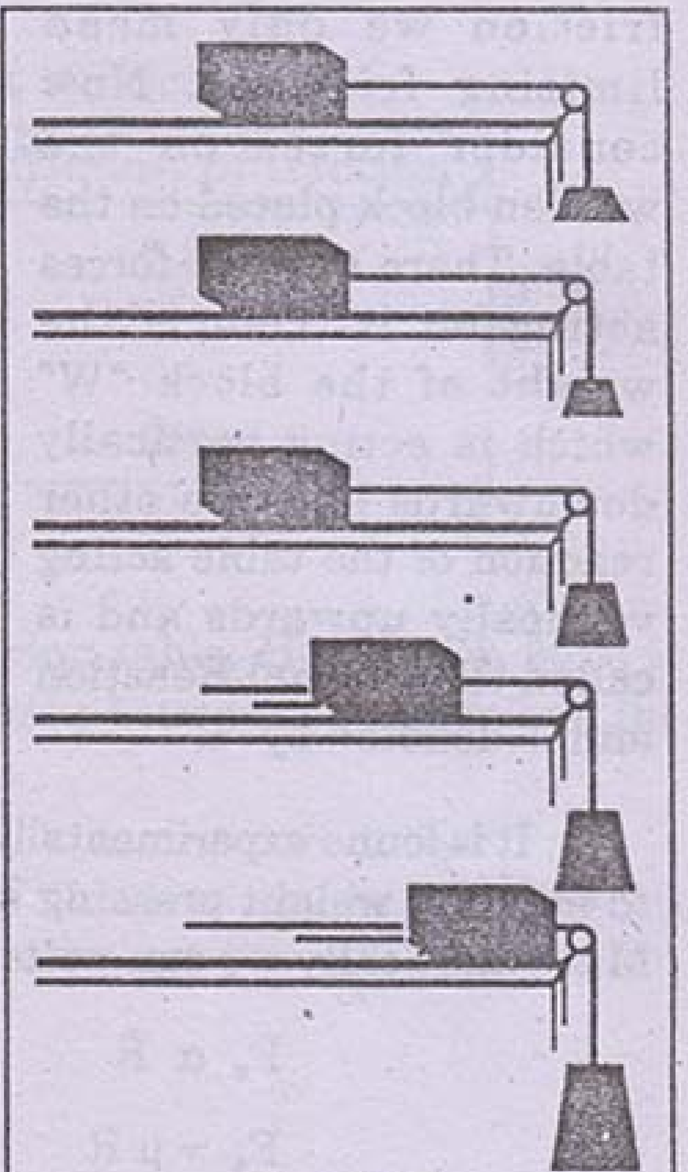


Fig. 4.12

We can observe the effect of friction by placing a wooden block on the horizontal surface of the table. The block is attached to one end of a string which passes over a pulley. A slotted weight is tied to the other end of the string.

A force "F" equal to the weight attached to the string acts on the block towards the right. Although this force is trying to move the block but is not sufficient to produce the motion in the block. The reason is that the opposing force of friction balances the applied force due to the weight.

Now go on increasing the weight gradually. You will see that the block does not move. As the applied force increases the force of friction also increases and reaches a maximum or limiting value which depends upon the nature of the surfaces in contact as well as the magnitude of the weight of the block and the block is on the point of sliding.

From the above observation we can conclude that force of friction is self adjusting force.

The maximum force of friction which just stops the body from sliding or moving is called limiting friction. After the motion starts there is no effect on the force of friction i.e. it remains constant after reaching its maximum or limiting value.

When a block is sliding over a surface, the frictional force which exists during the motion of the block is called the sliding friction and when a body e.g. a spherical ball rolls on a surface it experiences an opposing force called the rolling friction.

When we talk of friction we only mean limiting frictions. Now consider forces on the wooden block placed on the table. There are two forces acting on it. One is the weight of the block "W" which is acting vertically downwards and the other reaction of the table acting vertically upwards and is called the Normal Reaction and is denoted by "R".

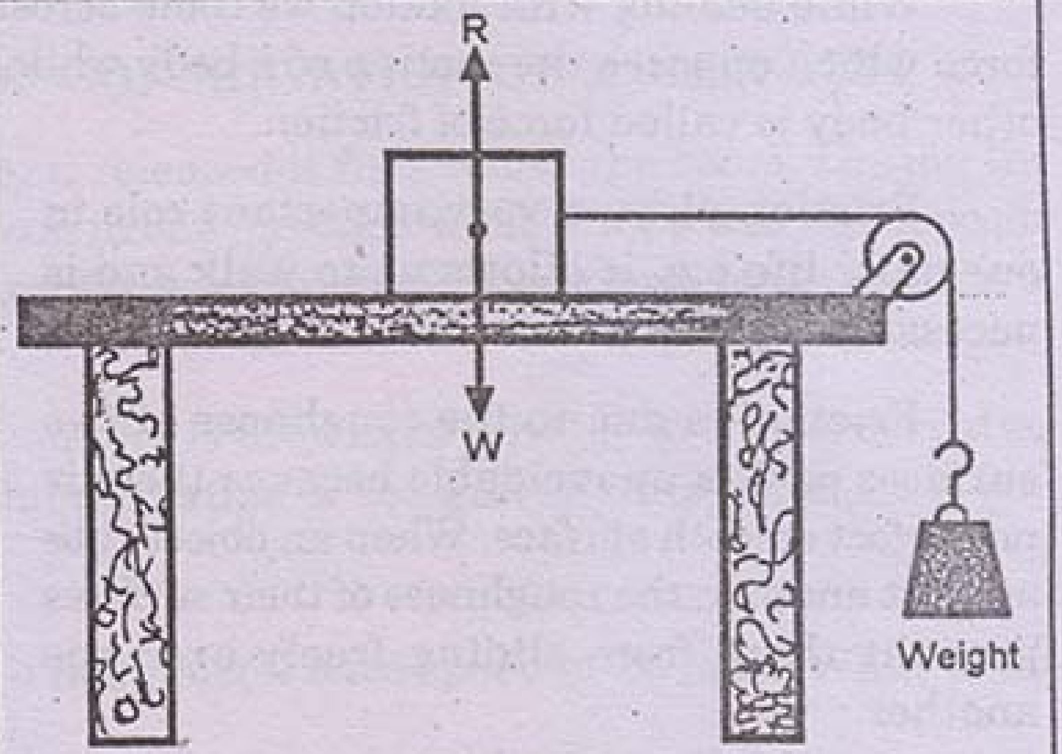


Fig. 4.13

It is found experimentally that the limiting frictional force "F_s" is proportional to the total weight pressing the block against the table or normal reaction "R". Mathematically we can write.

$$F_s \propto R$$

$$F_s = \mu R$$

$$\text{or} \quad \mu = \frac{F_s}{R} \longrightarrow (4.16)$$

$$\mu = \frac{\text{limiting frictional force}}{\text{Normal Reaction between the two surface in contact}}$$

Where "μ" is the constant of proportionately which depends upon the nature of the surfaces of the block and table and is called the coefficient of friction or simply coefficient friction.

As normal reaction is equal to the weight of the block if the surface is horizontal as shown in figure, therefore.

$$R = mg$$

Putting this value in equation 4.16 we get

$$\mu = \frac{F_s}{mg}$$

or $$\boxed{F_s = \mu \cdot mg}$$

Coefficient of friction is constant for a given pair of surfaces and is different for different pairs of surfaces. Values of μ for some pair of surfaces is given in Table 4.1

Table 4.1 Approximate Coefficient of Friction

Object (Dry Surface)	Co-efficient of friction
Between wood and wood	0.3
Between wood and stone	0.4
Between metal and metal	0.2
Between wood and leather	0.4

Example 4.5

A body weighing 50 N is placed on a wooden table. How much force is required to set it into motion? Coefficient of friction between the table and the body is 0.3.

Solution:

Weight of the body = $W = 50\ N$

Coefficient of friction = $\mu = 0.3$

Force = $F = ?$

$$F = \mu R$$

Since normal reaction $R = W$

$$F = \mu W$$

$$= 0.3 \times 50$$

$$F = 15\ N$$

Advantages and disadvantages of Friction:

Friction plays a vital role in our daily life. It has both advantages as well as disadvantages. Some times we have to increase the friction and some times we have to reduce the friction.

Advantages of Friction:

Friction is desirable for a number of purposes and in these cases we have to increase the friction. For example, sand is thrown on the up hill railway track after rain.

The threading on the car tyres is done to provide necessary frictional force between the tyre and the road to enable the car to move smoothly.

Friction enables us to walk on the ground.

Disadvantages of Friction:

The main disadvantage of friction is that a large amount of energy in machines is wasted due to it, which have to do much useless work in overcoming it. Moreover, friction leads to wear and tear on moving parts of machines.

Methods of Reducing Friction:

There are two different ways of reducing friction.

(a) The use of ball or roller-bearings.
(b) The use of lubricants like oil, air and graphite.

(a) Use of ball or roller-bearing:

In machines the sliding of various parts is usually replaced by rolling and this is done by using a ball bearing. Rolling friction is less than the sliding friction.

(b) Use of Lubricants:

The various parts of the machines which are moving over one another are properly lubricated.

The friction in different sliding surfaces can be reduced by putting fine chalk powder or any other powder or some liquid (oil) between them. The presence of dust or liquid keeps the sliding surfaces separated from each other and thus lessens the friction between them. This is the reason why oil, grease and air are used as lubricants.

Air under high pressure is used as lubricant in tools used for precision grindings.

SUMMARY

1. **The First Law of Motion:** Every body continues in its state of rest or of uniform motion in a straight line unless it is compelled to change that state by an external force impressed upon it.

2. **Law of Inertia:** The first law of motion is also called as law of Inertia. Inertia is that characteristic of a body due to which it resists against any change in its state of rest or of uniform motion.

3. **Force:** Force is that agent which changes or tends to change the state of rest or of uniform motion of a body. Force produces acceleration. It is a vector quantity and its unit is newton.

4. **The Second Law of Motion:** When an unbalanced force acts upon a body, it produces acceleration in the direction of the force and the magnitude of acceleration is directly proportional to the force and is inversely proportional to the mass.

5. **The Third Law of Motion:** To every action there is an equal but opposite reaction. Action and reaction do not act on the same body but act on two different bodies.

6. **Mass:** Mass of a body determines the magnitude of acceleration produced when a force acts on it. Mass does not change. It is a scalar quantity and its unit is kilogram.

7. **Weight:** The force with which Earth attracts a body towards its centre is called weight. The weight of a body can be different at different places. It is a vector quantity and its unit is newton.

8. **Tension in a String:** The force acting along a string is called tension.

9. **Motion of Bodies Attached to a String:** The motion of bodies attached to the two ends of a string passing over a pully can be of two types:

   (a) When one body moves vertically downward and the other upward, then the formulae of acceleration of bodies and tension in the string will be

   $$a = \frac{(m_1 - m_2)}{(m_1 + m_2)}, \quad T = \frac{2m_1m_2g}{(m_1 + m_2)}$$

   (b) When one body moves vertically downward and the other moves horizontally, then the formulae of acceleration and tension will be

   $$a = \frac{m_1 g}{(m_1 + m_2)}, \quad T = \frac{m_1m_2g}{(m_1 + m_2)}$$

10. **Momentum:** The product of mass and velocity of a moving body is called momentum. It is a vector quantity and its unit is $kgms^{-1}$ or Ns.

11. **Law of Conservation of Momentum:** When two or more than two bodies collide with one another, their total momentum remains the same provided no external force acts upon them.

12. **Friction:** The force that produces resistance against relative motion between two surfaces is called friction.

13. **Static Friction:** The resisting force between the two surfaces before the motion starts is called the static friction. The maximum value of the static friction is called as limiting friction.

14. **Kinetic Friction:** The friction during motion is called kinetic friction. This friction is slightly less than the limiting friction.

15. **Rolling Friction:** When a body moves with the help of wheels, the friction in this case is known as rolling friction. Rolling friction is much less as compared to the sliding friction.

QUESTIONS

4.1 Answer to the following questions.

(i) State Newton's three laws of motion, giving example from every day life.

(ii) Define force. How does second law of motion help in the measurement of force?

(iii) Explain why it is dangerous to jump from a fast moving train?

(iv) Explain the force of action and reaction in the following:
(a) Bullet is fired from a gun (b) A person sitting on a chair
(c) Motion of moon around the earth

(v) Differentiate between mass and weight.

(vi) Define momentum. Explain the law of conservation of momentum with the help of example.

(vii) Describe briefly the main cause of friction. Give three methods of reducing friction.

(viii) Why is the rolling friction less than the sliding friction?

4.2 Fill in the blanks.

(i) Motion cannot be produced in a body without __________.

(ii) No moving object can be stopped without applying __________.

(iii) The property of the matter by virtue of which it resists any attempt to change its state of rest or of uniform motion is called __________.

(iv) When an external force acts upon a body then it produces an __________ in the body in its own direction.

(v) The acceleration produced in a body under the influence of an external force is __________ proportional to the magnitude of the force.

(vi) The quantity of matter in a body is called its __________.

(vii) The force with which earth attracts a body towards its centre is called __________ of the body.

(viii) Every action has a reaction, these are __________ in magnitude but __________ in direction.

(ix) The product of mass and velocity is called __________.

(x) When one body slides over the surface of another body then unevenness of the surfaces result in __________ that causes the obstruction in the motion of the body.

(xi) Rolling friction is much __________ than sliding friction.

4.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) By applying equal force on spheres of plastic and iron, of equal volumes, greater acceleration is produced in the plastic sphere because its mass is __________.
(a) more (b) less (c) more but μ is less (d) less but μ is more

(ii) The SI unit of force is __________.
(a) metre (b) ms^{-1} (c) kg (d) newton.

(iii) The unit of coefficient of friction is __________.
(a) newton (b) kilogram (c) metre (d) none

(iv) Friction can be reduced by using ball bearings because they __________.
(a) make the surface plane
(b) make the surface grassy
(c) convert sliding friction into rolling friction
(d) have no friction of their own.

(v) If the force acting on a body is doubled, then the acceleration produced is. ______.

(a) 1/2 (b) 1/4 (c) double (d) quadrupled

(vi) When a horse pulls a wagon, the force that causes the horse to move forward is the force.

(a) he exerts on the wagon (b) the ground exerts on him
(c) the wagon exerts on him (d) the wagon exerts on the ground

(vii) Which is the best approximation of the weight of an object of mass 800 gram?

(a) 88N (b) 80N (c) 8N (d) 0.8N (e) 0.08N (f) 7.840N

4.4 Pick out true and false from the following sentences.

(i) Newton's first law of motion is also known as law of inertia.

(ii) The acceleration produced in a body under the influence of an external force, is inversely proportional to the applied force.

(iii) Mass is a vector quantity where as weight is a scalar quantity.

(iv) Friction is a self adjusting force which is equal to but in the opposite direction of the applied force just before the motion begins.

(v) The maximum force of friction is called limiting friction.

(vi) The value of the limiting friction is inversely proportional to normal reaction.

(vii) Had there been no friction it would not have been possible to set a body in motion or to stop a moving body.

(viii) The biggest disadvantage of friction is that some energy is continually being lost due to friction between some parts of machines.

(ix) The value of the coefficient of friction depends upon the nature of the surfaces in contact.

(x) It is wrong to say that mass is the measure of inertia.

PROBLEMS

4.1 Determine the acceleration of a car of mass 900 kg, when a net force of 2700 N acts on it.

($3ms^{-2}$)

4.2 A car of mass 1000 kg travelling at 72 kmh^{-1} is uniformly brought to rest over a distance of 40 m. Find (a) the average acceleration; (b) the average breaking force.

($-5ms^{-2}$, -5000 N)

4.3 A bullet of mass 50 g travelling with a speed of $15ms^{-1}$ penetrates into a bag of sand and is uniformly brought to rest in 0.05s.

Find (a) How for the bullet will penetrate into the bag of sand?
(b) The average force exerted by the sand?

($-300ms^{-2}$, -15 N)

4.4 A force of 120 N acts on a stationary body for 4 seconds and the body acquires a velocity of $36ms^{-1}$. Calculate the mass of the body.

(13.3 kg)

4.5 A gun of mass 20 kg fires a bullet of mass 50 g with a velocity of 200 ms^{-1}. Calculate the velocity of the recoil of the gun.

(0.5 ms^{-1})

4.6 An empty truck weighs 4000 N. Its engine can produce a maximum acceleration of 1 ms^{-2}. If the truck is loaded with 2000N, find the maximum acceleration the engine can produce.

(0.66 ms^{-2})

4.7 Two bodies A and B are attached to the end of a string which passes over a pulley, so that they hang vertically. If the mass of the body B is 4 kg; find mass of the body A which moves up with an acceleration of 0.5 ms^{-2}. (Take g = 10 ms^{-2})

(3.48 Kg)

4.8 A rectangular metal block of mass 4 kg rests on the top of a metal surface. The coefficient of friction between the box and the metal surface is 0.2. What force parallel to the surface is needed to move the block.

(8 N)

CHAPTER – 5

VECTORS

LEARNING OBJECTIVES:

- Introduction.
- Vector representation.
- Multiplication of a vector by a number.
- Negative of a vector
- Trigonometric ratio.
- Resolution of vectors.

5.1 INTRODUCTION:

We deal with a number of physical quantities in physics, such as force, mass, acceleration, momentum, energy, etc. Some of these quantities are said to be scalars. Such quantities only have a magnitude and a unit of measurement, for example, time, distance, energy, volume, speed, etc. The normal rules of arithmetic are used to add, subtract, multiply and divide scalar quantities.

Other quantities, such as force, displacement, momentum, and velocity etc., have a magnitude, a unit of measurement and a direction. These quantities are called vectors quantities. Vector quantities are added, subtracted and multiplied by special rules.

The physical quantities which are completely specified by their magnitude and direction both, are called vector quantities. For example, if a drill-master asks a boy to move three steps, the boy does not know in which direction he has to move. If, however, he mentions the direction also, the confusion is removed. Hence displacement is a vector quantity. The distance moved by the boy and the direction of his motion give the magnitude and the direction of the displacement vector respectively. Thus displacement is a vector quantity. Force is another example of a vector; when we push or pull a body, we are said to

exert a force on it. To describe a force, we need to specify its direction, as well as its magnitude. Velocity, acceleration etc., are also vectors as none of these quantities give a complete meaning without the mention of direction.

There are many ways of denoting vector quantities. In this book, a vector will be denoted by a letter with an arrow over it and its magnitude by the same letter without an arrow vector. The vector $\vec{A}$ has magnitude A. Vectors are added, subtracted, multiplied by the rules of vector algebra.

5.2 VECTOR REPRESENTATION:

A vector is represented by an arrow drawn parallel to the direction of the vector. The length of the arrow, on a suitable scale, indicates the magnitude of the vector, and the arrow-head gives its direction.

Suppose a person moves a distance of 50 kilometre towards north from a certain reference point. To represent this displacement vector we first indicate north-south direction (Fig. 5.1)

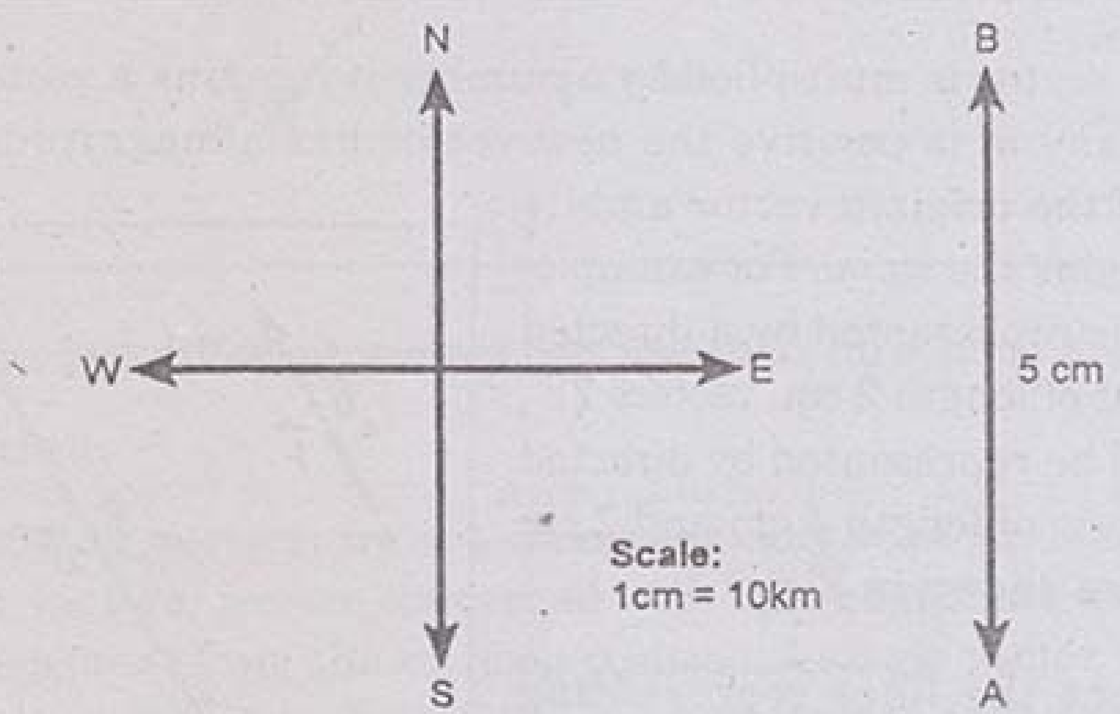


Fig. 5.1: A displacement of 50 kilometre due north is represented by the directed line segment $\vec{AB}$.

Secondly we select a suitable scale, say a distance of 10 kilometre is represented by 1.0 centimetre. A directed line segment $\vec{AB}$, 5 cm in length, drawn parallel to north direction, represents the displacement vector of 50 Km towards north. A similar procedure is adopted to represent all other types of vectors.

In the above example the direction of the vector has been indicated with respect to north-south and east-west directions represented by two lines drawn at right angles to each other. However, in some cases a vector is more conveniently represented with respect to any two convenient lines drawn mutually at right angle. One of these lines is named as x-axis and the other as y-axis. Such two

lines are known as reference axes or rectangular axes. For example, if a boat is moving with a velocity of 10 km/h in a direction making an angle of 30° with the bank of the river, the velocity of the boat can be conveniently represented by considering the x-axis parallel to the bank of the river, as shown in Fig. 5.2.

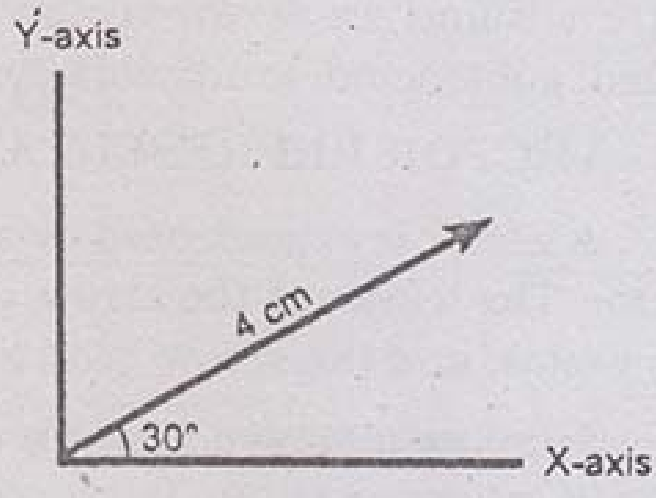


Fig. 5.2: A line segment making an angle of 30° with x-axis gives magnitude and direction of velocity of the boat by considering the x-axis parallel to the bank of the river.

5.3 MULTIPLICATION OF A VECTOR BY A NUMBER:

When a vector is multiplied by a number it remains a vector quantity. If the number, say n, is positive the new vector has a magnitude n times the magnitude of the original vector and its direction remains the same. For example, if a vector $\vec{F}$ is represented by a directed F line segment of length 2 cm, vectors $2\vec{F}$, and $\frac{1}{2}\vec{F}$ will be represented by directed lines segments of length 4 cm and 1 cm respectively, as shown in Fig. 5.3.

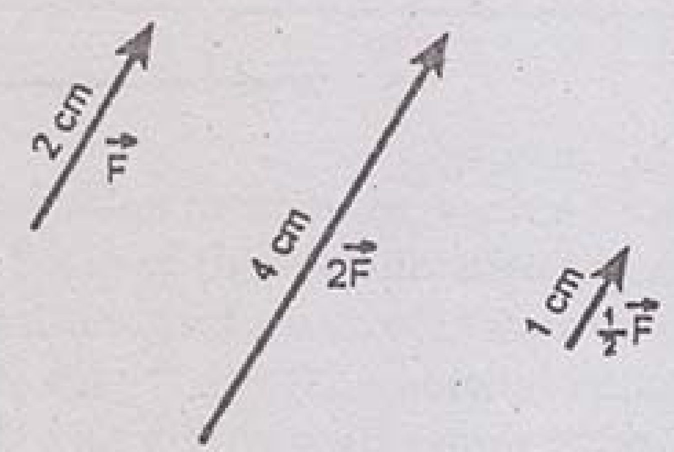


Fig. 5.3

Note that the lines representing these vectors are parallel to one another, with their arrow-heads pointing in the same direction.

In case the vector $\vec{F}$ is multiplied by a negative number, say −n, the new vector has a magnitude n times the magnitude of the original vector and direction, opposite to that of $\vec{F}$. The vectors $\vec{F}$, $-2\vec{F}$ and $-\frac{1}{2}\vec{F}$ have been shown in Fig. 5.4. Note that the negative sign causes a reversal of the direction of the arrow-head.

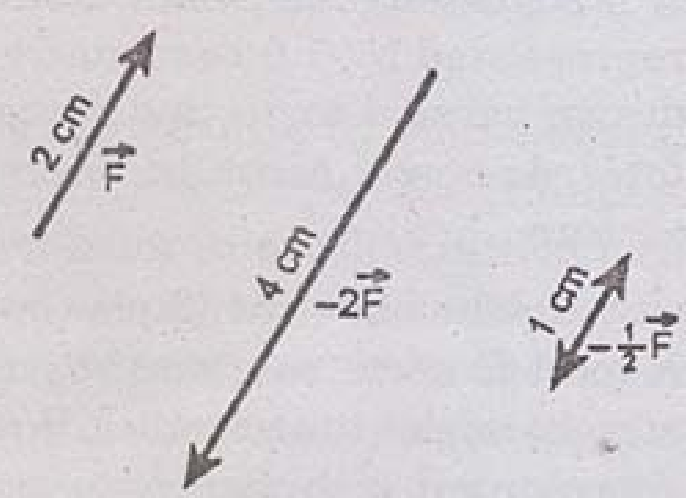


Fig. 5.4

5.4 NEGATIVE OF A VECTOR:

If we reverse the direction of a given vector without changing its magnitude or multiply a vector by −1, it becomes a negative of the vector. For example, the given vector is represented in Fig. 5.5 by $\overrightarrow{OA}$. The negative of this vector is represented in the same Fig. by $\overrightarrow{OB}$.

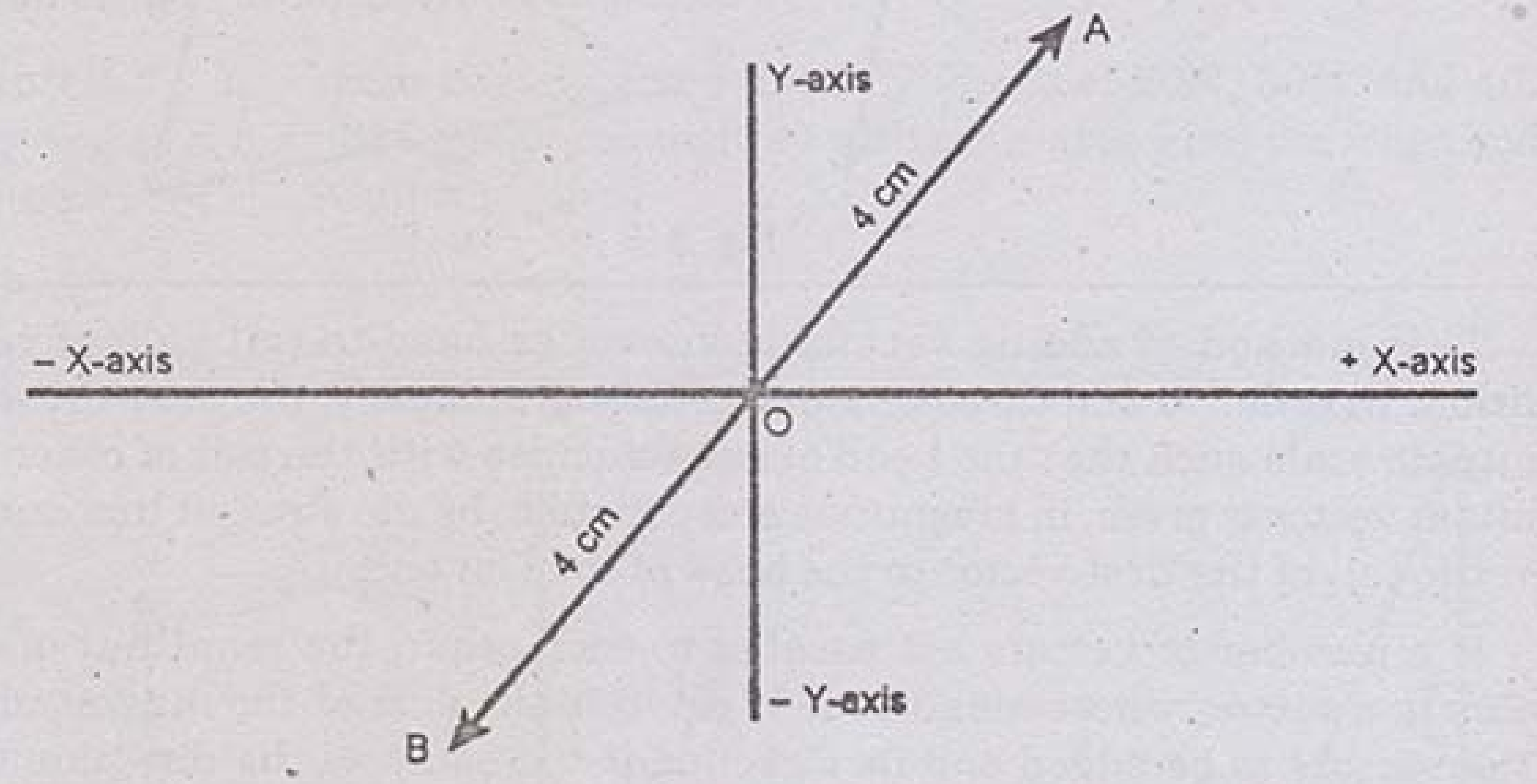


Fig. 5.5

These are two different vectors because their directions are different.

Vector addition:

When adding vectors, we are concerned only with their net result. For displacement vectors, we are concerned with the magnitude and direction of our net displacement from the starting position. A single vector whose effect is the same as the combined effect of the given vectors is called the resultant of the vectors. Suppose we want to add two vectors $\vec{A}$ and $\vec{B}$. Their resultant $\vec{R}$ is given by:

$$\vec{R} = \vec{A} + \vec{B}$$

Graphical Method:

To add them graphically, we use the following method. Draw the representative lines of vectors $\vec{A}$ and $\vec{B}$ on a suitable scale with respect to a system of convenient reference axes. Redraw their representative lines such that the head of $\vec{A}$ coincides with the tail of $\vec{B}$. Draw the representative lins of another vector $\vec{R}$, from the tail of $\vec{A}$ to the head of $\vec{B}$, as shown in Fig. 5.6. This vector $\vec{R}$ represents the resultant of $\vec{A}$ and $\vec{B}$.

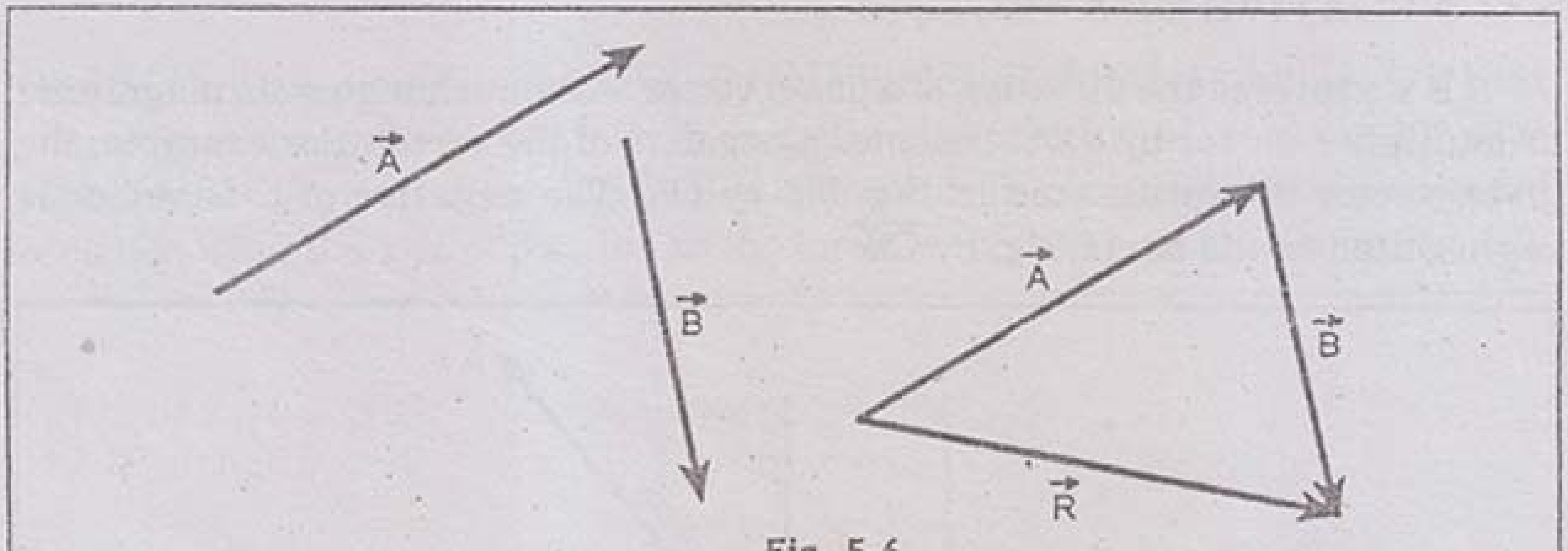


Fig. 5.6

This method of adding vectors is known as head-to-tail rule of vector addition. In order to add three or more vectors graphically, they are drawn on a suitable scale such that the head of one coincides with the tail of other. The resultant vector is given, in magnitude and direction, by the straight line directed from the tail of the first vector to the head of the last one.

If a number of vectors act parallel to each other, the resultant of such vectors is a vector, whose magnitude is equal to the sum of the magnitudes of all the vectors to be added and its direction is the same as the direction of all the vectors.

Example 5.1: A car travels 200 km west, 100 km south, and finally 150 km at an angle 60° south of east, what is the net displacement of the car?

Solution:

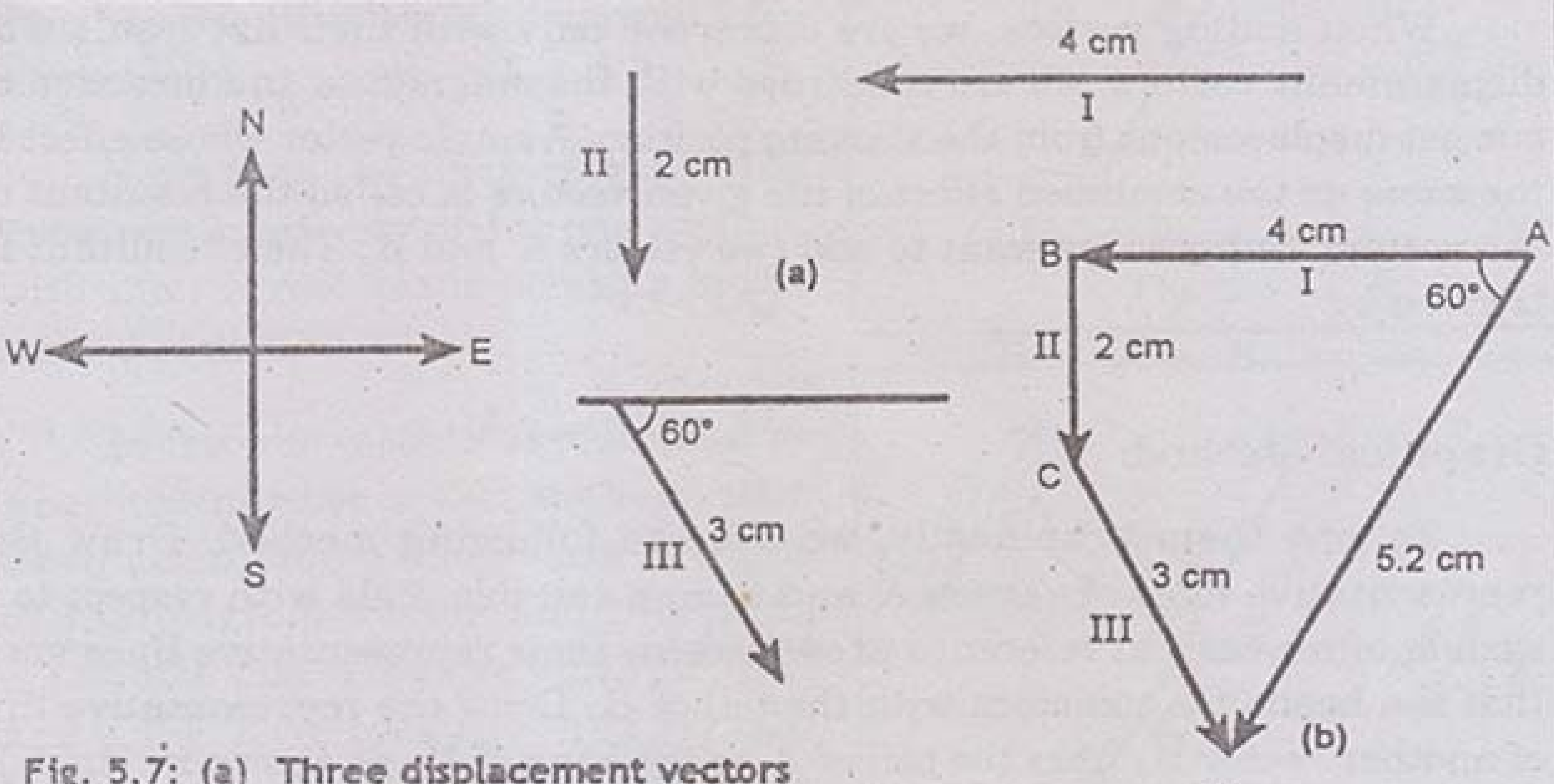


Fig. 5.7: (a) Three displacement vectors
(b) The resultant displacement

The given displacement vectors are drawn by choosing a suitable scale, shown in Fig. 5.7(a). The resultant displacement $\overrightarrow{R}$ is obtained with the help of the head-to-tail rule of vector addition, which is shown in Fig. 5.7(b). The length of the resultant displacement $\overrightarrow{AD}$ is measured, which comes out to be 5.2 cm. Since one centimetre represents 50 Km., the magnitude of the resultant displacement is 5.2×50 Km = 260 Km. Using a protractor, we find that the direction of the resultant is 60° south of west.

Example 5.2: A certain body is acted upon by forces of 20N, 30N, and 40N making angles of 30°, 60° and 90° respectively with the x-axis. Find the magnitude and direction of the resultant force?

Solution:

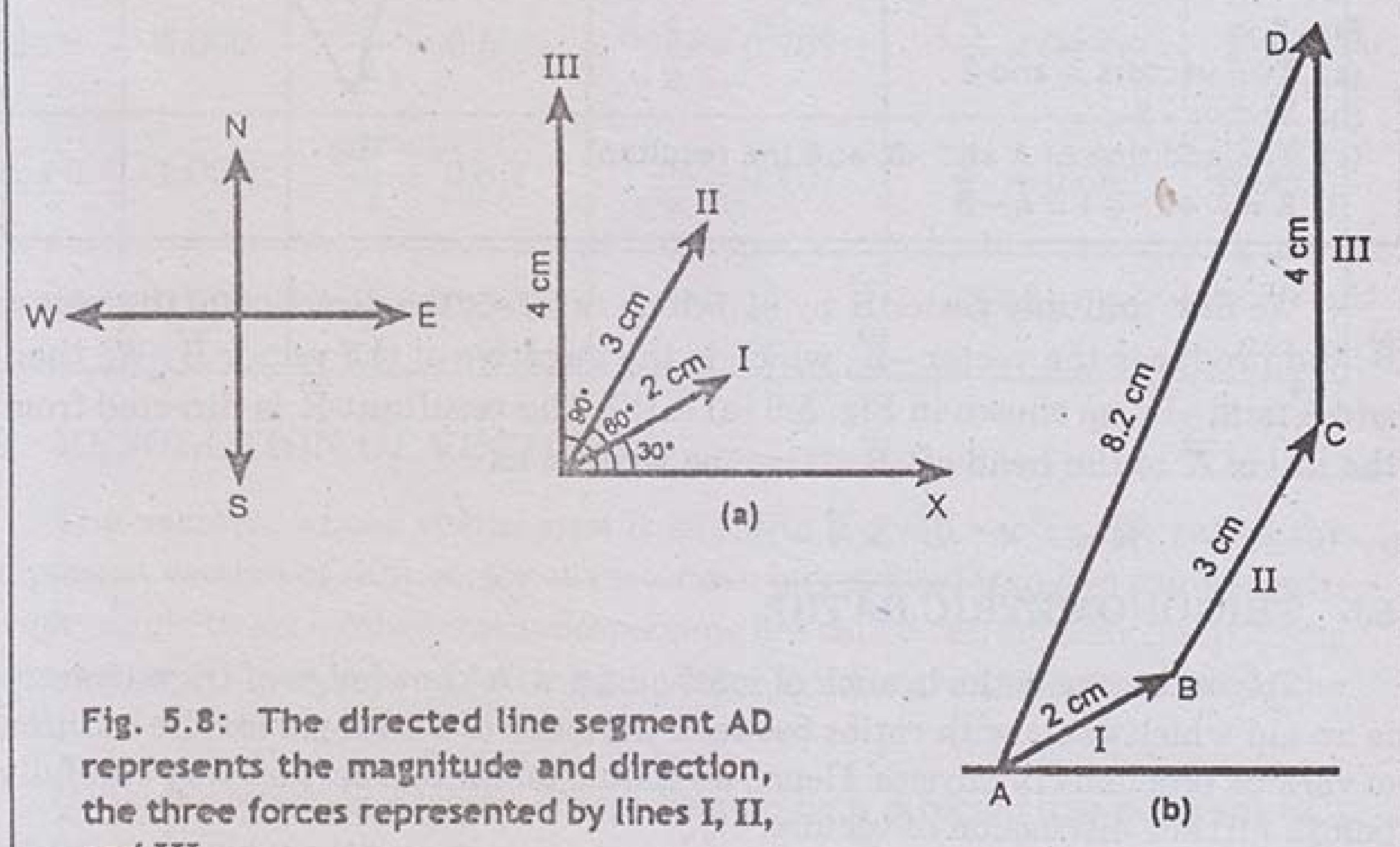


Fig. 5.8: The directed line segment AD represents the magnitude and direction, the three forces represented by lines I, II, and III.

The three given forces are represented in magnitude and direction by directed line segments I, II and III (Fig. 5.8a). The resultant of these forces, as determined by head-to-tail rule, is represented in magnitude and direction by directed line segment $\overrightarrow{AD}$. On measurement, the length of $\overrightarrow{AD}$ comes out to be 8.2 cm and the angle which it makes with the x-axis is 67°. Thus the magnitude of the resultant force is 82 N and it acts in a direction making an angle of 67° with x-axis.

Vector Subtraction:

Consider the subtraction of vectors. Let a vector $\vec{B}$ to be subtracted from vector $\vec{A}$.

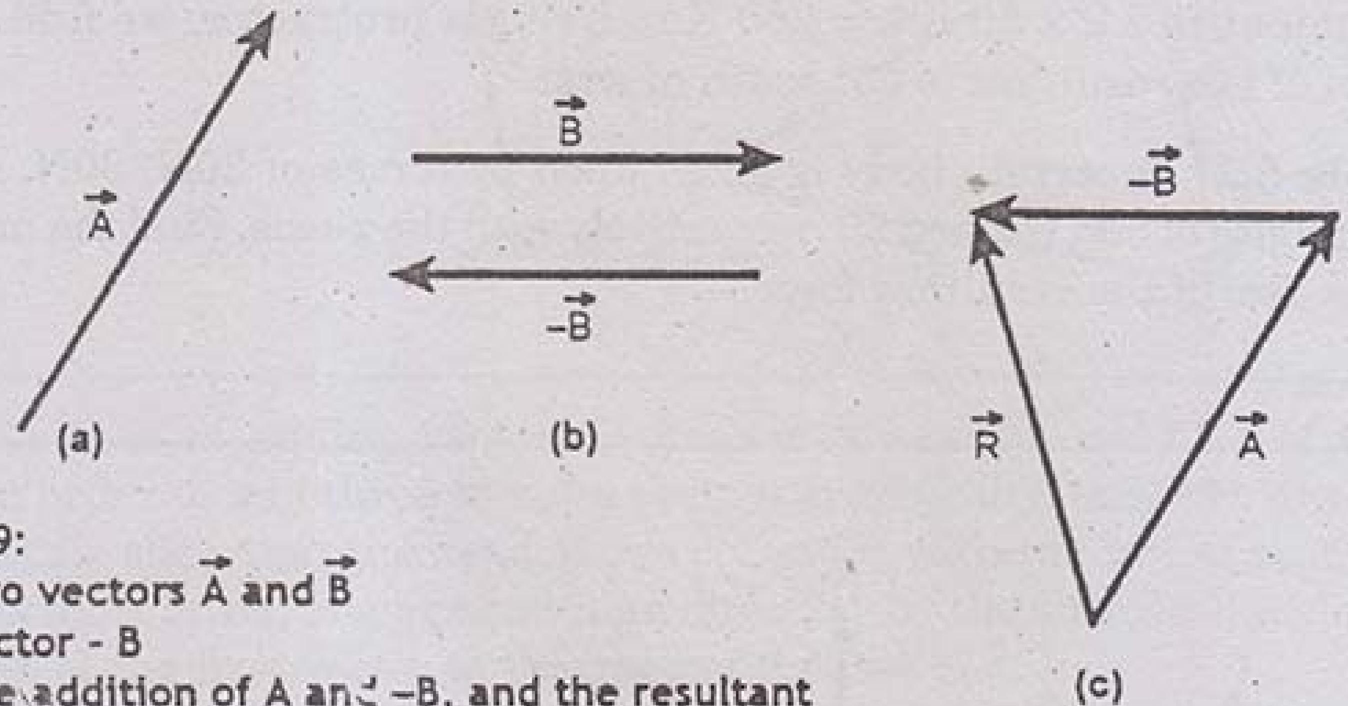


Fig. 5.9:
(a) Two vectors $\vec{A}$ and $\vec{B}$
(b) Vector - B
(c) The addition of A and −B, and the resultant $\vec{R} = \vec{A} + (-\vec{B}) = \vec{A} - \vec{B}$

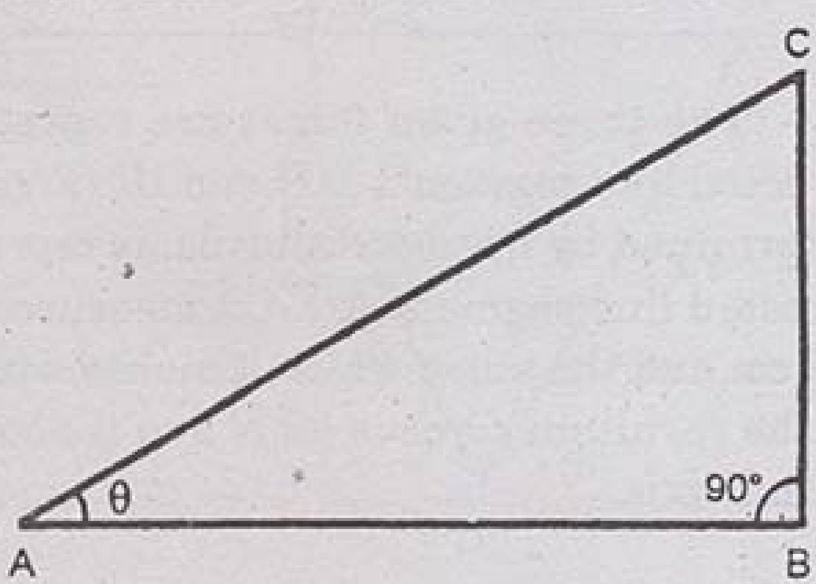
We first multiply vector $\vec{B}$ by −1, which reverses the direction of the vector $\vec{B}$ and produces the vector $-\vec{B}$, which is the negative of the vector $\vec{B}$. We then add $\vec{A}$ and $-\vec{B}$, as shown in Fig. 5.9 (a) & (b). The resultant $\vec{R}$ is directed from the tail of $\vec{A}$ to the head of $-\vec{B}$. It can be written as

$$\vec{R} = \vec{A} + (-\vec{B}) = \vec{A} - \vec{B}$$

5.5 TRIGONOMETRIC RATIO:

Trigonometry is the branch of mathematics. A knowledge of trigonometry is an aid which deals with ratios between the sides of a triangle for the solution of various problems in physics. Hence we should follow this section very carefully before further discussion of vectors.

We consider a right angled triangle ABC in which BC and AB are opposite and adjacent sides to the angle θ and are generally called as perpendicular and base, respectively. The side opposite to right angle is called hypotenuse. We have different trigonometric ratios which give the ratios between the various sides of a right angled triangle. They are



$$\sin\theta = \frac{\text{Opposite Side}}{\text{Hypotenuse}} = \frac{BC}{AC}$$

$$\cos\theta = \frac{\text{Adjacent Side}}{\text{Hypotenuse}} = \frac{AB}{AC}$$

$$\tan\theta = \frac{\text{Opposite Side}}{\text{Adjacent Side}} = \frac{BC}{AB}$$

Trigonometric ratios of some common angles are given in Table 5.1.

Table 5.1 Trigonometric Ratios

	0°	30°	45°	60°	90°
$\sin\theta$	0.000	$\frac{1}{2} = 0.500$	$\frac{1}{\sqrt{2}} = 0.707$	$\frac{\sqrt{3}}{2} = 0.866$	1.000
$\cos\theta$	1.000	$\frac{\sqrt{3}}{2} = 0.866$	$\frac{1}{\sqrt{2}} = 0.707$	$\frac{1}{2} = 0.500$	0.000
$\tan\theta$	0.000	$\frac{1}{\sqrt{3}} = 0.577$	1.000	$\sqrt{3} = 1.732$	$\propto$

5.6 RESOLUTION OF VECTORS:

The vectors, whose vector sum is equal to a given vector, are called the component vectors of that vector. A vector can be resolved into two components, at right angle to each other, such components are called rectangular components of the vector. The process of splitting a vector into its components is called the resolution of the vector.

Consider a vector $\vec{F}$ which is represented by a directed line segment $\overrightarrow{OC}$, making an angle θ with x-axis.

From the terminal point C of the vector $\vec{F}$ we draw perpendiculars CA and CB on the x- and y-axes respectively. Now consider two vectors represented by directed line segments $\overrightarrow{OA}$ and $\overrightarrow{OB}$ or $\overrightarrow{AC}$ as shown in Fig. 5.10. As the direction of $\overrightarrow{OA}$ is along x-axis, we shall denote this vector by $\vec{F_x}$. Similarly the vector $\overrightarrow{OB}$ or $\overrightarrow{AC}$, having its direction along or parallel to y-axis, is denoted by the vector $\vec{F_y}$. By applying the head to tail rule of vector addition, it can be seen in the Fig. 5.10 that the sum of the vectors $\vec{F_x}$ and $\vec{F_y}$ is the vector $\vec{F}$. Hence $\vec{F_x}$ and $\vec{F_y}$ are the components of vector $\vec{F}$. As the angle between $\vec{F_x}$ and $\vec{Fy}$ is 90° hence they are called rectangular components. The magnitude of these components

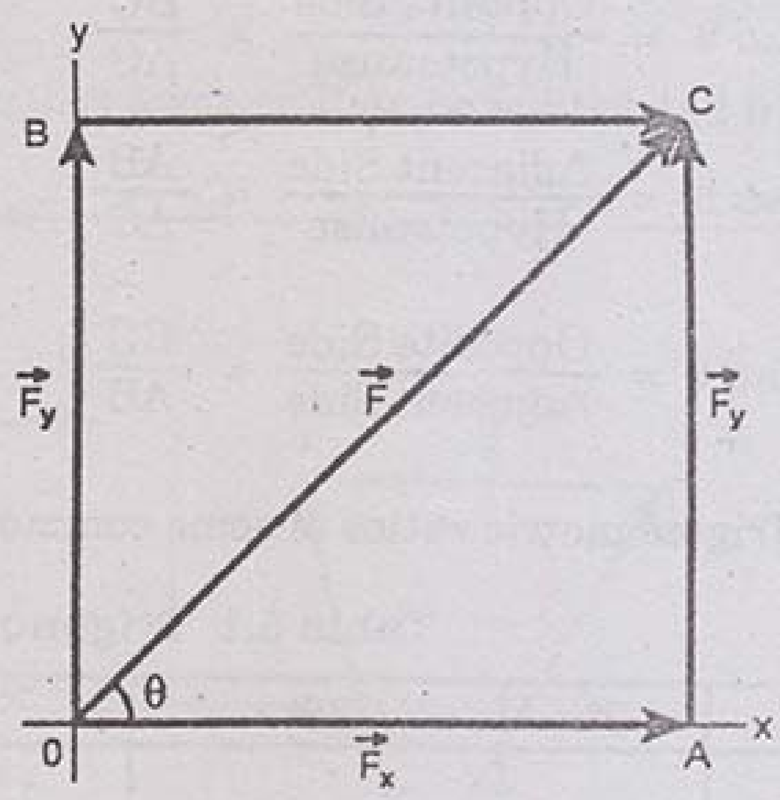


Fig. 5.10:
Vectors $\vec{F_x}$ and $\vec{F_y}$ are rectangular components of $\vec{F}$ in the directions of x- and y-axis.

can be determined in terms of magnitude of $\vec{F}$ and the trigonometric ratios of angle θ. In the right angled triangle OAC, we have

$$\cos\theta = \frac{OA}{OC}$$

$$\therefore OA = OC\cos\theta$$

As the lengths of the directed line segments $\overrightarrow{OC}$ and $\overrightarrow{OA}$ represent the magnitude of the vectors $\vec{F}$ and $\vec{F_x}$ respectively, so

$$F_x = F\cos\theta \quad \text{——————} (5.1)$$

Similarly

$$\sin\theta = \frac{AC}{OC}$$

$$\therefore AC = OC\sin\theta$$

$$\therefore F_y = F\sin\theta \quad \text{——————} (5.2)$$

Thus rectangular components of a vector can be determined by using equations (5.1) and (5.2).

Composition of a vector by rectangular components:

If the rectangular components of a vector are given, the magnitude and direction of the vector can be obtained. Thus a vector can be completely specified by its two rectangular components.

Consider rectangular components $\overrightarrow{F_x}$ and $\overrightarrow{F_y}$ which are represented by directed line segments $\overrightarrow{OA}$ and $\overrightarrow{OB}$ respectively, both magnitude and direction. Adding these two components by head-to-tail rule, it can be seen, from Fig. 5.11 that the directed line segment $\overrightarrow{OC}$ represents, in both magnitude and direction of the vector $\vec{F}$.

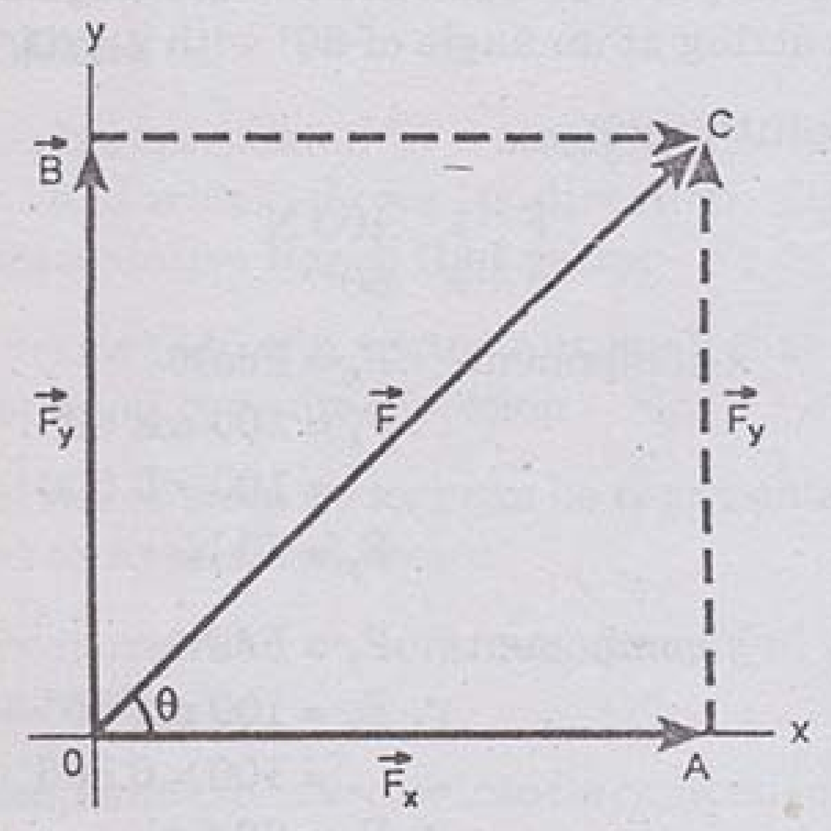


Fig. 5.11: $\vec{F}$ is vector sum of rectangular components $\vec{F}_x$ and $\vec{F}_y$.

To derive the expressions for the magnitude and direction of $\vec{F}$ in terms of the magnitudes of the rectangular components $\vec{F}_x$ and $\vec{F}_y$, consider the right angled triangle OAC. By using Pythagorus theorem, we have

$$OC^2 = OA^2 + AC^2$$

$$\text{or} \quad F^2 = F_x^2 + F_y^2$$

$$\text{or} \quad F = \sqrt{F_x^2 + F_y^2} \quad \text{———} \quad (5.3)$$

The direction of $\vec{F}$ is determined by finding the angle θ which the vector $\vec{F}$ makes with the x-axis. In the right angled triangle OAC, we have

$$\tan\theta = \frac{AC}{OC}$$

$$\text{or} \quad \tan\theta = \frac{F_y}{F_x} \quad \text{———} \quad (5.4)$$

$$\therefore \quad \theta = \tan^{-1}\frac{F_y}{F_x}$$

Thus the equations (5.3) and (5.4) completely determine the vector $\vec{F}$ in terms of its two rectangular components.

Example 5.3: Find the two rectangular components of a force of 100 N which is acting at an angle of 60° with x-axis.

Solution:

$$F = 100\text{ N}$$
$$\theta = 60^\circ$$

x-component, $F_x = F\cos\theta$

$$\therefore F_x = 100\cos 60^\circ$$
$$= 100 \times 0.500$$
$$F_x = 50\text{ N}$$

y-component, $F_y = F\sin\theta$

$$\therefore F_y = 100\sin 60^\circ$$
$$= 100 \times 0.866$$
$$\therefore F_y = 86.6\text{ N}$$

Example 5.4: The x and y-components of a force vector $\vec{F}$ are 3N and 4N respectively; find the magnitude and direction of $\vec{F}$.

Solution:

$$F_x = 3\text{ N}$$
$$F_y = 4\text{ N}$$
$$F = \sqrt{F_x^2 + F_y^2}$$
$$F = \sqrt{(3)^2 + (4)^2}$$
$$= \sqrt{9 + 16}$$
$$= \sqrt{25}$$
$$\therefore F = 5\text{ N}$$

And $\tan\theta = \frac{F_y}{F_x} = \frac{4}{3}$

$$\therefore \tan\theta = 1.333$$
$$\theta = \tan^{-1}(1.333) = 53^\circ \text{ approx}$$

The angle θ can be determined by consulting the table of natural trigonometric ratios for different angles or by using scientific calculation.

SUMMARY

1. **Vector Representation:** A vector is represented by a straight line of a specific length having an arrow head which shows its direction. This straight line is also called the representative line of that vector.

2. **Negative Vector:** The negative vector (–A) of a vector **A** is such that it has the same magnitude as that of A but opposite direction.

3. **Resultant Vector:** The addition of two or more vectors can be represented by a single vector, which is termed as a resultant vector.

4. **Addition or Subtraction of Vectors:** The vectors may be added or subtracted by using the head to tail rule.

5. **Resolution of a Vector:** The decomposition of a vector into its components is called resolution of a vector.

6. **Rectangular Components:** The components of a vector which are perpendicular to each other. The magnitude of the components are

$$F_x = F\cos\ ,\ F_y = F\sin\theta$$

7. **To find vector from its rectangular components:** If the rectangular components $\mathbf{F_x}$ and $\mathbf{F_y}$ are given, then the magnitude of the vector F is

$$F = \sqrt{F_x^2 + F_y^2}$$

and its direction

$$\theta = \tan^{-1}\frac{F_y}{F_x}$$

QUESTIONS:

5.1 Write answers to the questions given below:

1. Explain scalar and vector quantities.
2. Mention which of the following quantities are scalars and which are vectors. Speed, velocity, force, mass, weight, displacement, work, temperature, wavelength, torque, acceleration, momentum.
3. How can a vector be represented in magnitude and direction both?
4. What do you know about multiplication of a vector by a number?
5. Define the terms negative vector and resultant vector.
6. Explain head to tail rule of vector addition.
7. What are rectangular components of a vector? How are they determined?
8. How can a vector be determined if its rectangular components are known?
9. Is it possible to combine two vectors of different magnitude to give a zero resultant? If not, can three vectors be so combined?
10. Can a force directed north balance a force directed east? Explain.

5.2 Fill in the blanks.

1. A __________ has only magnitude.
2. A __________ has magnitude as well as direction.
3. A vector is represented by a straight line in such a way that its length indicates __________ and arrow-head indicates __________ of the vector.
4. Vectors are added or subtracted graphically by __________ rule.
5. The angle between rectangular components of a vector is __________.
6. The x- and y-components of a force $\vec{F}$ acting at an angle θ with the x-axis are given by

   $F_x =$ ______________, $F_y =$ ______________.
7. In a right angled triangle $\sin\theta =$ ______, $\cos\theta =$ ______, $\tan\theta =$ ________.
8. If Fx and Fy are rectangular components of a force $\vec{F}$, then $F = \sqrt{\text{.......} + \text{.......}}$

5.3 Given below are a few possible answers to each statement. Identify the correct one.

1. ________ is a scalar quantity.

   (a) Torque. (b) Distance.
   (c) Momentum. (d) Acceleration.

2. ________ is a vector quantity.
   (a) Work. (b) Density.
   (c) Velocity. (d) Temperature.

3. If a force $\vec{F}$ is multiplied by a number n, the magnitude of the new vector becomes ____________.
   (a) n+F (b) n–F
   (c) nF (d) $\frac{F}{n}$

4. The unit of force in International system of units is _____
   (a) Kilogram. (b) newton.
   (c) Metre. (d) Second.

5. In a right angled triangle the side opposite to the right angle is called __________.
   (a) Hypotenuse. (b) Perpendicular.
   (c) Base. (d) None of these.

6. In a right angled triangle sinθ is equal to the ratio of ________.
   (a) Base to perpendicular.
   (b) Perpendicular to base.
   (c) Hypotenuse to perpendicular.
   (d) Perpendicular to hypotenuse.

7. If F_x and F_y are rectangular components of a force $\vec{F}$, then tanθ = ________.
   (a) $\frac{F_x}{F_y}$ (b) $\frac{F_y}{F_x}$
   (c) $F_x + F_y$ (d) $F_x - F_y$

5.4 Choose true and false from the following sentences.

1. Acceleration is a vector quantity.
2. Cosθ =
3. Vectors can be added by head to tail rule.
4. The process of splitting up a single vector into two vectors is called composition of a vector.
5. Two components, at right angle to each other, are called rectangular components of a vector.

PROBLEMS:

5.1 Draw the representative lines of the following vectors.

(i) Force of 15N making an angle of 60° with x-axis.
(ii) Displacement of 60 Km in the direction of north.
(iii) Velocity of 30 Km/h towards north-east.

5.2 A body moves 5Km towards east from a fixed point A and reaches a point B. From B it covers 3Km towards north and arrives at a point C. Find the distance and direction of the displacement.

(Ans. 5.8Km; 31° north of east)

5.3 A man drives his car 5Km east of his starting point. Then he travels 10Km south and reaches a point. From here he travels 15Km towards west and reaches another point. Finally he travels 20Km towards north. Find the net displacement of the man from where he started.

(Ans. 14.2Km; 45° west of north)

5.4 Three forces of magnitude 3N, 4N and 4N are acting at an angle of 0°, 30° and 150° with the x-axis respectively. Find the resultant force.

(Ans. 5N; 53° with the x-axis)

5.5 Find the x and y-components of the following forces by trigonometric method.

(i) N at 45° with the x-axis.
(ii) 15N at 60° with the x-axis.
(iii) 20N at 90° with the x-axis.

(Ans. x-components 10N, 7.5N, 0
y-components 10N, 12.99N, 20N)

5.6 Determine completely the resultant displacement vector in each of the following cases:

	x-components	y-components
(i)	5Km	3Km
(ii)	7.5Km	18Km
(iii)	5.3Km	43.8Km

(Ans. 5.38Km, 31°; 19.5Km, 67°; 44Km, 83°)

CHAPTER – 6

EQUILIBRIUM

LEARNING OBJECTIVES:

- Introduction.
- Parallel forces.
- Torque or moment of a force.
- Centre of gravity.
- Couple.
- Equilibrium of bodies under the action of coplanar forces.
- Conditions of equilibrium.
- States of Equilibrium.

6.1 INTRODUCTION:

A body that is either at rest or moving with uniform speed in a straight line or rotating with a uniform speed is said to be in equilibrium. The study of forces acting on stationary bodies in equilibrium is called statics. A body at rest is in static equilibrium and that moving in a straight line with uniform speed is in dynamic equilibrium. In this chapter we shall study what conditions are necessary for a body to remain in equilibrium. In order to obtain equilibrium, we must see that no unbalanced force is applied to the body. Generally a number of forces act on a body but they are balanced to give a resultant force equal to zero. Sometimes the resultant force on a body is zero, but the body may still have a rotational acceleration about its axis, so for a body to remain in equilibrium, it must satisfy two conditions. Each of these conditions can be written in the form of a mathematical equation. These equations will be discussed in this chapter and are used to analyze a variety of situations.

6.2 PARALLEL FORCES:

When a number of forces act on a body and if their directions are parallel, they are called parallel forces. Any two given parallel forces may be like or

unlike. If two parallel forces have the same direction, they are called like parallel forces. If two parallel forces have opposite directions, they are called unlike parallel forces.

(a) Resultant of two like parallel forces.

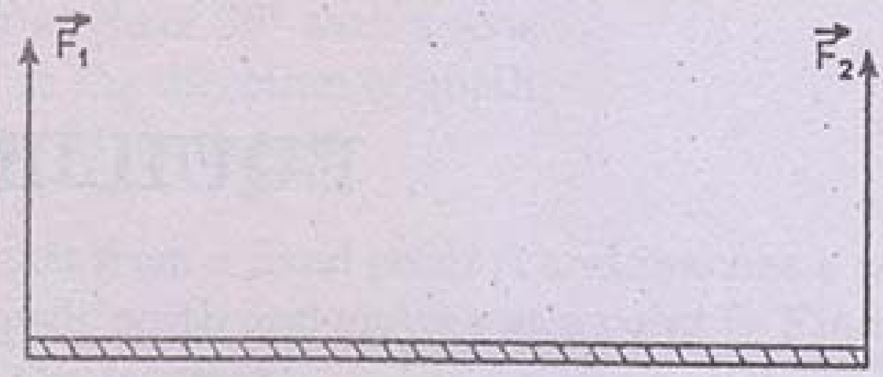


Fig. 6.1: Two like parallel forces acting on a body.

Consider two like parallel forces $\vec{F}_1$ and $\vec{F}_2$ acting on a body at A and B, as shown in Fig. 6.1. Suppose $\vec{R}$ is the resultant force of $\vec{F}_1$ and $\vec{F}_2$, then

$$R = F_1 + F_2$$

It means that the resultant of two like parallel forces is a force whose magnitude is equal to the sum of the magnitudes of the two forces and the direction is the same as either of the forces.

(b) Resultant of two unlike parallel forces.

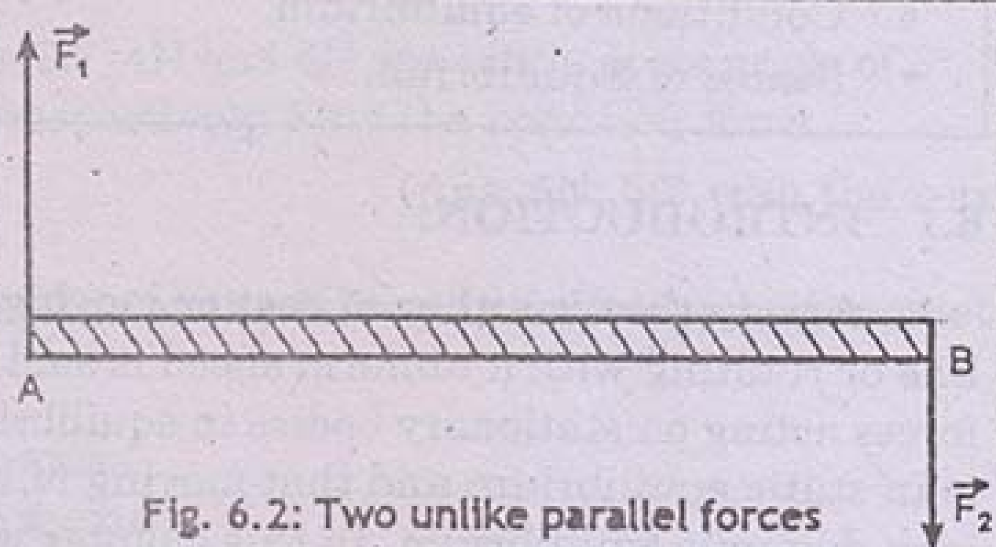


Fig. 6.2: Two unlike parallel forces acting on a body.

Consider two unlike parallel forces $\vec{F}_1$ and $\vec{F}_2$ acting on a body at A and B, as shown in Fig. 6.2. Suppose $\vec{R}$ is the resultant force of $\vec{F}_1$ and $\vec{F}_2$. Here F_1 is greater than F_2. Thus

$$R = F_1 - F_2$$

It means that the resultant of two unlike parallel forces is also a force whose magnitude is equal to the difference of the magnitudes of the two forces and the direction is the same as that of the greater force.

Axis of rotation:

Some bodies can not move from place to place but they can rotate about a fixed line or axis, called its axis of rotation. A force acting on such a body can only produce rotation. For example, when we apply a force on door, it can rotate, as it can not move as a whole along a straight line. Thus an axis, about which a body is free to rotate, is called axis of rotation.

6.3 TORQUE OR MOMENT OF A FORCE:

If a body is capable of rotating about an axis, then a force applied properly

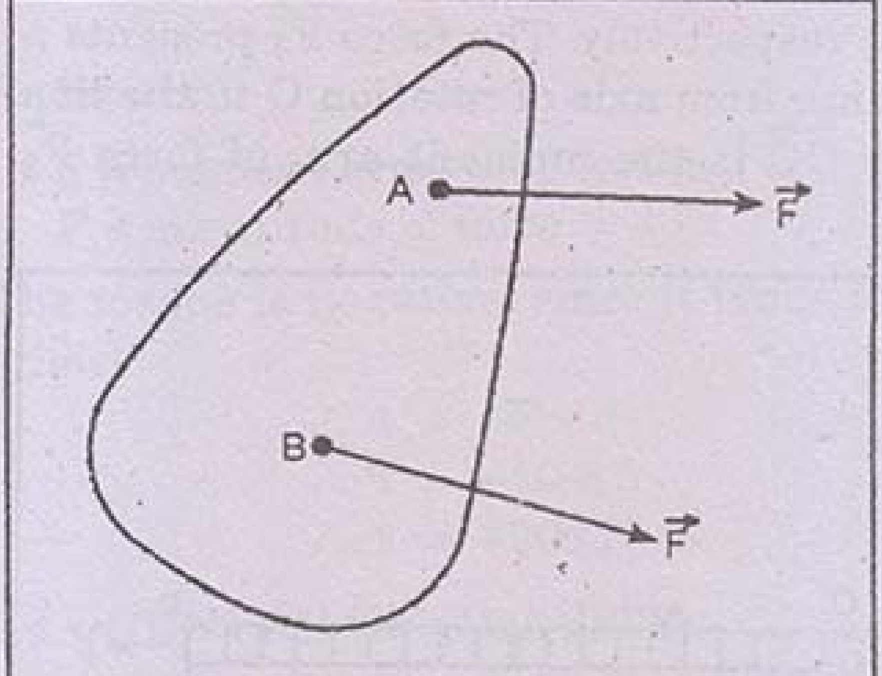


Fig. 6.3: The turning effect produced by a force depends upon the position of the line of action of the force

on this body will rotate it about that axis. This turning effect of the force is known as torque or moment of the force. The turning effect produced by a force of a given magnitude and direction depends upon the position of the line of action of the force. A body will rotate clockwise if a force $\vec{F}$ is applied at a point A, whereas it will rotate counter-clockwise when the same force is applied at B, (Fig. 6.3)

We shall study the factors that determine the turning effect about some axis when a number of co-planar forces (forces acting in the same plane) act on a body. Consider the rotation of a heavy wheel (Fig. 6.4). It is a common experience that the wheel can be rotated more easily and quickly by applying a force $\vec{F}$, perpendicular to a spoke at some point A, far from the axis, than by applying the same force $\vec{F}$ at a point B, nearer the axis. This means that the turning effect of a force is greater the farther is the line of action of the force from the axis of rotation. The distance of the line of action of the force from the axis is measured perpendicular distance to the line of action of the force. It is not the distance from the axis to the point of application of the force. If the same force $\vec{F}$ is applied at C rather than at A where OC = OA, there will be no turning effect as the line of action of $\vec{F}$ now passes through axis and so the perpendicular distance from axis of rotation to the line of action of force is zero. The perpendicular distance from axis of rotation to the line of action of the force is called the moment arm of the force. Thus one of factors that determine the torque or moment of a force is moment arm of the force.

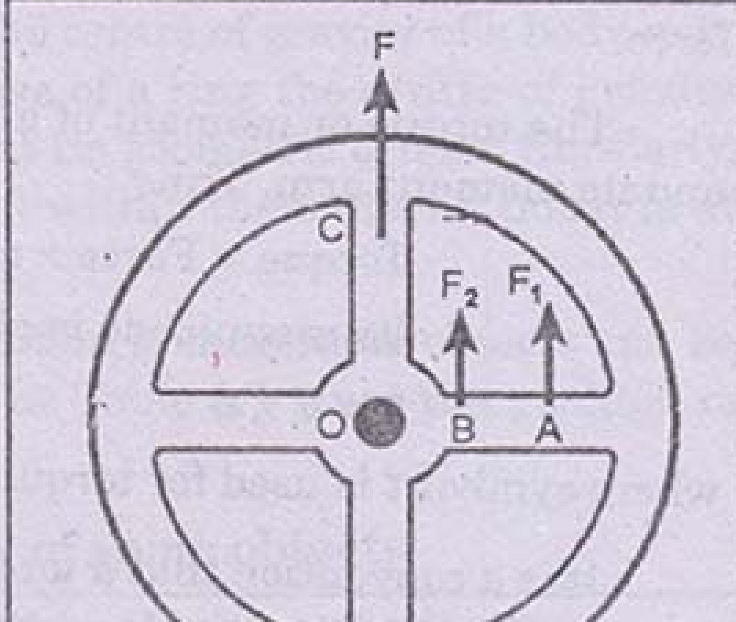


Fig. 6.4: The turning effect of a force is greater, the farther is the line of action of the force from the axis of rotation.

In order to study the moment arms of the forces acting on a bar AB suppose the bar is capable of rotation about its mid point O. Let the length of the bar be L as shown in Fig. 6.5. The moment arm of force $\vec{F}$ is zero as its line of action

passes through the axis at O. The forces $\vec{F_1}$ and $\vec{F_2}$ act perpendicular to the bar, so their moment arms are $\frac{L}{2}$ and $\frac{L}{4}$ respectively. The force $\vec{F_3}$ presents a different case. The perpendicular distance from axis of rotation O to the line of action of the force $\vec{F_3}$ is $\overline{OC}$. Thus $\overline{OC}$ is the moment arm of force $\vec{F_3}$ about O.

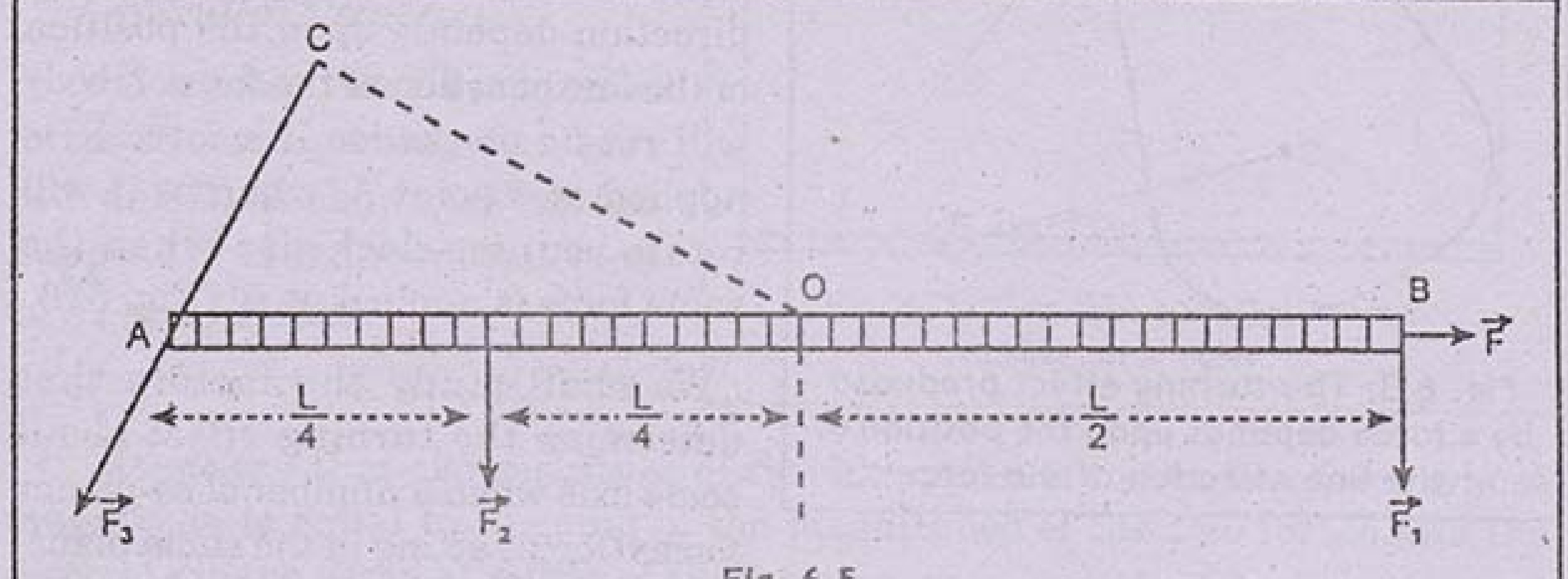


Fig. 6.5

Another factor that determines the torque or moment of a force is magnitude of the force. It can be observed that greater the force, the greater is torque or moment of a force. Thus the magnitudes of the force and the moment arm are two quantities which are important in measuring the torque or moment of a force.

The torque or moment of a force about an axis is the product of the force and its moment arm. Thus

Torque = Force × moment arm

its magnitude is given by

$$\tau = F \times d$$

where symbol τ is used for torque, F for force and d for moment arm of the force.

It is a convention that a torque which produces a counter-clockwise rotation is considered as positive and that produces a clockwise rotation is taken as negative. The unit of torque in S.I. units is newton-meter i.e. N-m.

Example 6.1: Calculate the torque produced by a force of 40N on the beam arranged as shown in Fig. 6.6.

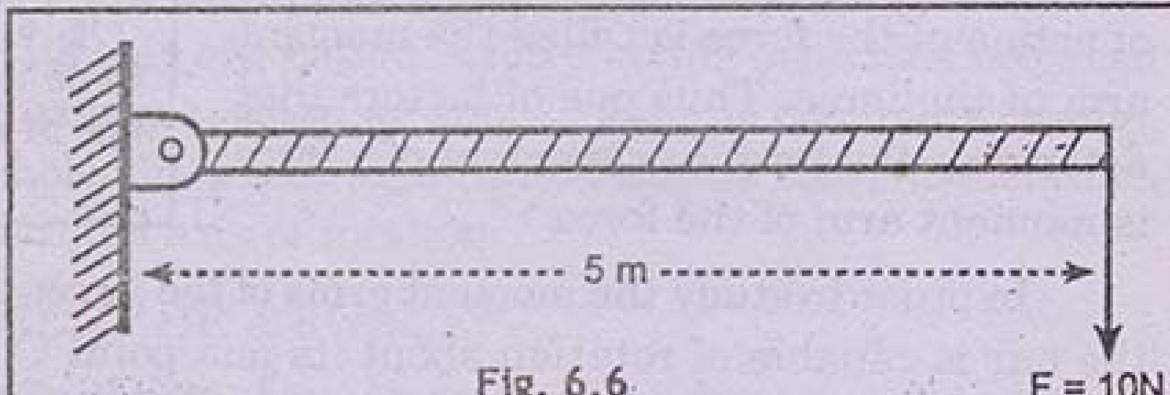


Fig. 6.6

Solution:

d = moment arm = 5 m

F = magnitude of force = 40N

The torque is negative, since it tends to rotate the beam clockwise about the origin.

$$\therefore \tau = -F \times d$$

$$\therefore \tau = -40 \times 5$$

$$\therefore \tau = -200 \text{ N-m}$$

6.4 CENTRE OF GRAVITY:

Every object is made of a number of particles, each of which is attracted by the earth with a force. The direction of the force on each particle is towards the centre of the earth. Distance of the particles from the centre of the earth is so great that all these forces can be considered parallel to one another. Hence the force of gravity acting on a body is the resultant of all these parallel forces. This force is called weight of the body. The weight of the body always acts vertically downward through a point, which is called centre of gravity of the body. Thus centre of gravity can be defined as that point on which the weight of the body acts.

The centre of gravity of a body depends on the shape of the body. In a sphere the centre of gravity is at its centre. The centre of gravity of a body may not necessarily be within the body. In the case of a ring the centre of gravity does not lie in the material of the ring. It is at its geometric centre. Similarly, in a hollow sphere the centre of gravity is not within the material but it is at its geometric centre.

The centre of gravity of regular-shaped and homogenous objects can be found from the consideration of symmetry. The Table 6.1 gives the position of the centre of gravity of some familiar bodies.

Table 6.1: Centre of gravity of some objects

Name of body	Position of centre of gravity
1. Uniform rod	Centre of rod
2. Circular plate	Centre of plate
3. Plate (Square, Rectangle or parallelogram in shape)	Intersection of lines joining mid-points of opposite sides.
4. Triangular plate	Intersection of medians
5. Rectangular block or cube	Intersection of diagonals
6. Cylinder	Mid-point of axis.

The knowledge of the position of centre of gravity is useful in problems of equilibrium, because that is the point of application of the vector representing the weight of the body.

Centre of gravity of an irregular shaped body:

Plumb line is a small metallic conical piece attached to a string. When the plumb line is displaced, it always comes to rest in the vertical position. Two vertical forces act on it, (1) its weight (2) tension in the string. These forces act along the same vertical line. Thus vertical line passes through the centre of gravity of the plumb line. The centre of gravity of the plumb line is vertically below the point of suspension. This fact provides a simple method for finding experimentally the centre of gravity of irregular shaped bodies. This method is as under:

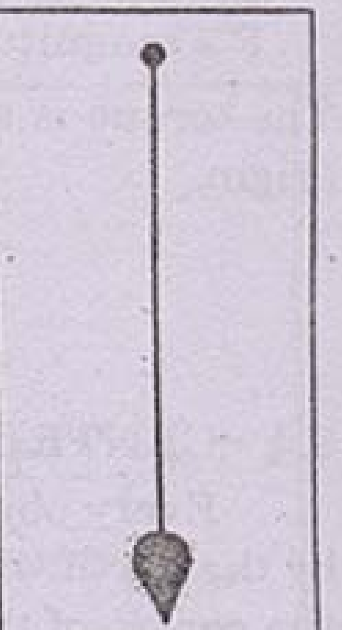


Fig. 6.7: Plumb line

Drill three small holes near the edge of the plate whose centre of gravity is to be located (Fig. 6.8). Suspend the plate from a nail fired horizontally in a wall using one of the holes, say A. When the plate is at rest, suspend a plumb line from the nail. Draw a line Aa on the plate along the plumb line. The centre of gravity lies somewhere on the line Aa.

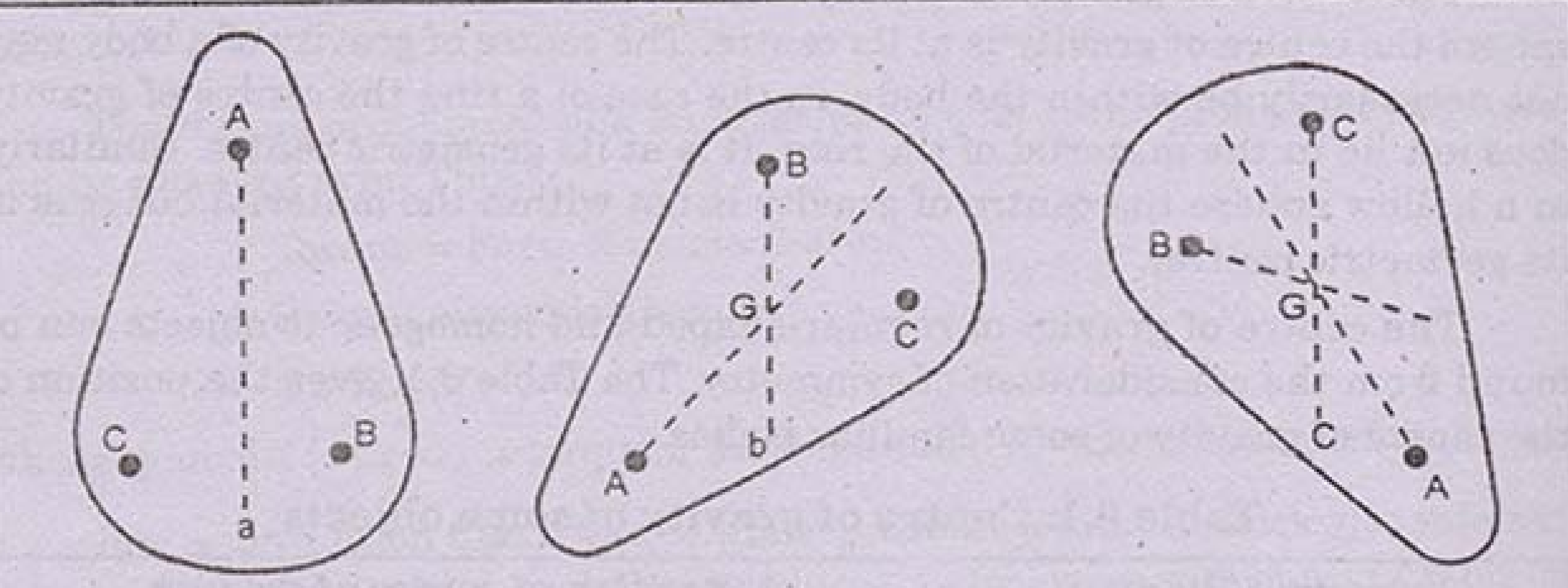


Fig. 6.8

Repeat the above procedure with hole B on the nail. Again the centre of gravity must lie somewhere on the line Bb. The only point common to the lines Aa and Bb is G, therefore this point must be the centre of gravity. If the plate is suspended using third hole C, the line Cc will also pass through the point G.

6.5 COUPLE:

A pair of equal, parallel and unlike forces having different lines of action, is called a couple.

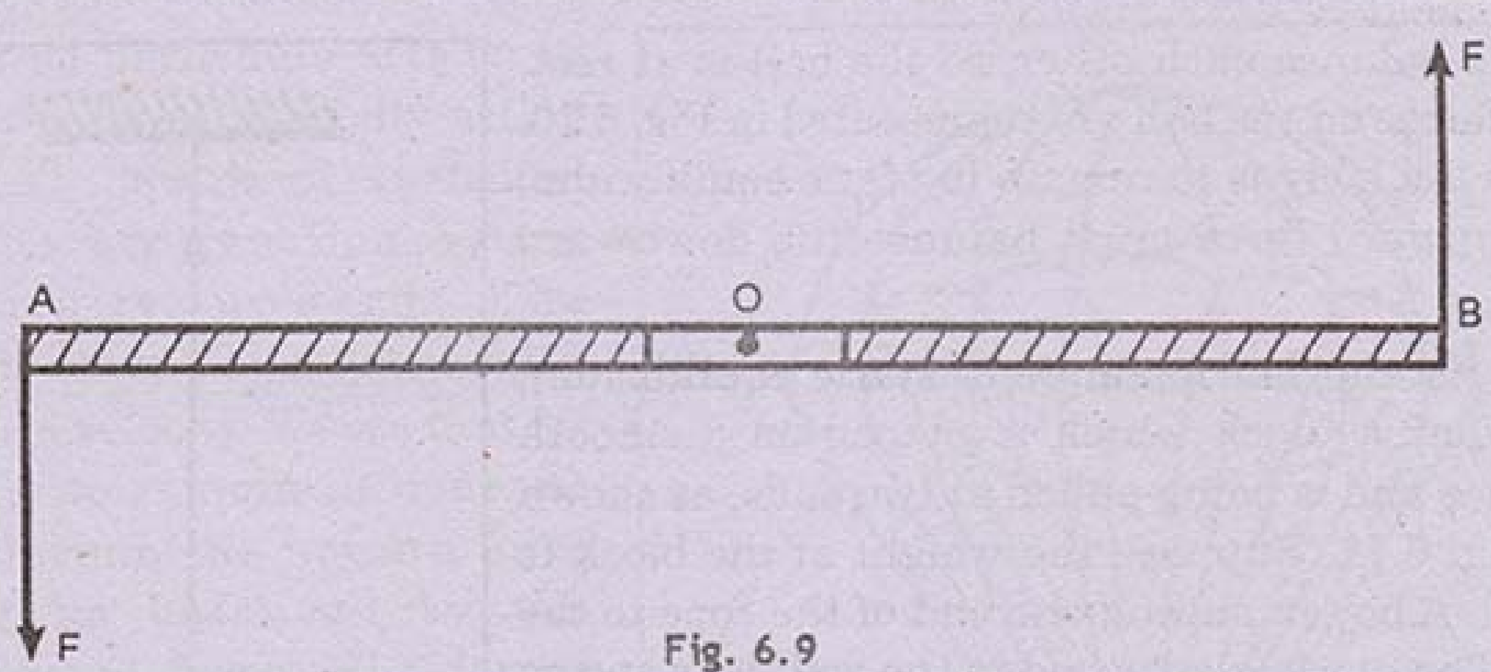


Fig. 6.9

Consider two equal, unlike parallel forces, each of magnitude F, acting at A and B, as shown in the Fig. 6.9. The torques or moments of two forces are given by

The moment of the force F at A = $F \times OA$
The moment of the force F at B = $F \times OB$

Both these moments have the same direction i.e. counter clockwise, so the total moment of the two forces is equal to the sum of the two moments.

$$\therefore \text{ Moment of the couple } = F \times OA + F \times OB$$
$$= F(OA + OB)$$
$$= F \times AB$$

Moment of the couple is equal to the product of one of the forces and the perpendicular distance between the lines of action of two forces. This perpendicular distance between the two forces is called the arm of the couple.

Thus the torque or moment of a couple is equal to the product of either force and arm of the couple. Whenever a couple acts on a body, there is rotation only. It should be noted that a couple cannot be balanced by a single force. It can be balanced by an equal and opposite couple.

6.6 EQUILIBRIUM OF BODIES UNDER THE ACTION OF COPLANAR FORCES:

If a body remains at rest or moves with uniform speed, it is said to be in equilibrium. A body at rest is said to be in static equilibrium, while a body in uniform motion along a straight line is said to be in dynamic equilibrium. All the bodies, in static or dynamic equilibrium, do not possess any acceleration.

Let us consider a spherical ball of weight 5N suspended from the ceiling by a string as shown in Fig. 6.10. The ball is in static equilibrium. There are two forces acting on the ball. (i) The force of gravity W = 5N directed downward, (ii) The upward pull of the string, known as tension T in the string. These two

forces balance each other, so the ball is at rest. The forces on the ball are represented in Fig. 6.10. Thus if a body is to remain in static equilibrium, the upward force must balance the downward force.

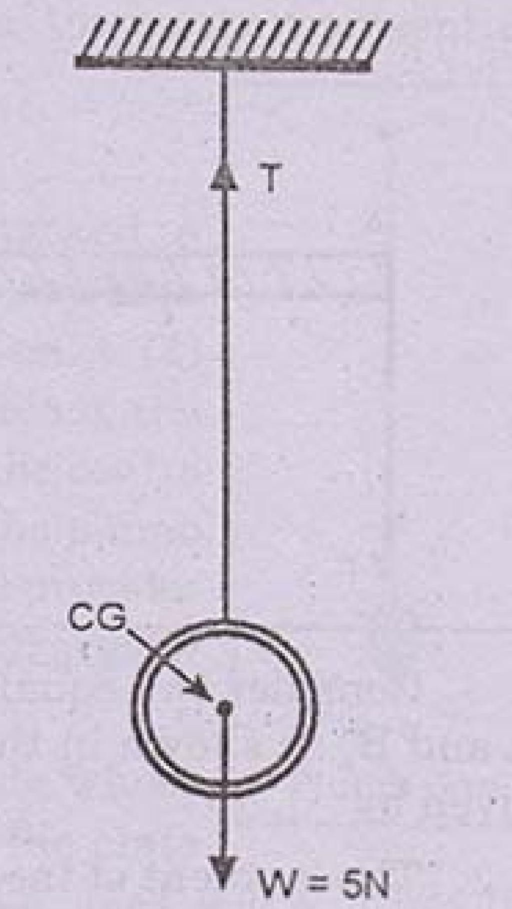


Fig. 6.10: As the ball is in equilibrium, the tension in string must balance the force of gravity.

As another example of static equilibrium, consider a block which is placed on a smooth surface and is being pulled at two ends, as shown in Fig. 6.11. Suppose the weight of the block is 100N. A boy is pulling one end of the rope to the left. The block is attached to the wall by a spring balance which reads 50N. For simplicity we ignore the force of friction. Four forces are acting on the block and it is at rest. Weight of the block acting vertically downward is balanced by the push of the surface acting vertically upward. The block is being pulled towards left by hand and it is being pulled towards right by the spring balance. As the block remains at rest, so these forces must also balanced each other. If the block is to remain in static equilibrium, the force towards left must be balanced by the force towards right and the upward force must be balanced by the downward force. The forces acting on the block are represented in Fig. 6.11. Thus the resultant of all the four forces acting on the block is zero.

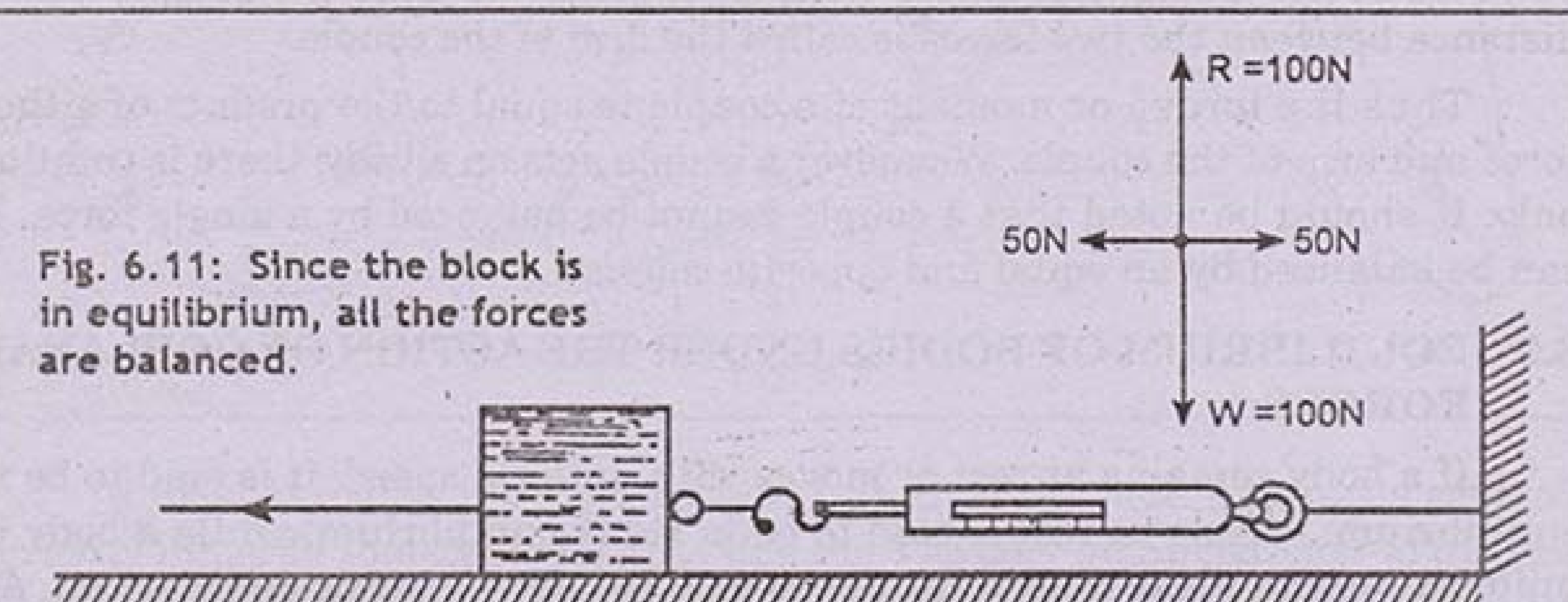


Fig. 6.11: Since the block is in equilibrium, all the forces are balanced.

A paratrooper falling with uniform velocity is an example of dynamic equilibrium. When a paratrooper jumps from an aeroplane, it moves down with an acceleration due to gravity. The force of gravity on the paratrooper acts downward. Thereafter, the parachute opens and an upward force due to reaction

of air on parachute acts on it. Now the force of gravity acting vertically downward is balanced by the air reaction acting vertically upward. The paratrooper falls down with uniform velocity. We see that for a body to remain in dynamic equilibrium the upward force must be balanced by the downward force. The forces acting on the paratrooper are represented in Fig. 6.12.

Fig. 6.12: After a parachute opens and falls a certain distance, it moves downward thereafter with uniform velocity.

Another example of dynamic equilibrium can be presented by considering a car moving on a road with uniform velocity (Fig. 6.13).

The force of engine is acting in the forward direction while the force of friction between road and tires is acting backward. If these two forces balance each other, then the car will move with uniform velocity.

It is clear from the above examples that for a body to be in equilibrium, no unbalanced force should act on the body. The downward forces must balance the upward forces and the leftward forces must balance the rightward forces. It also means that the resultant of all the forces acting on a body should be zero.

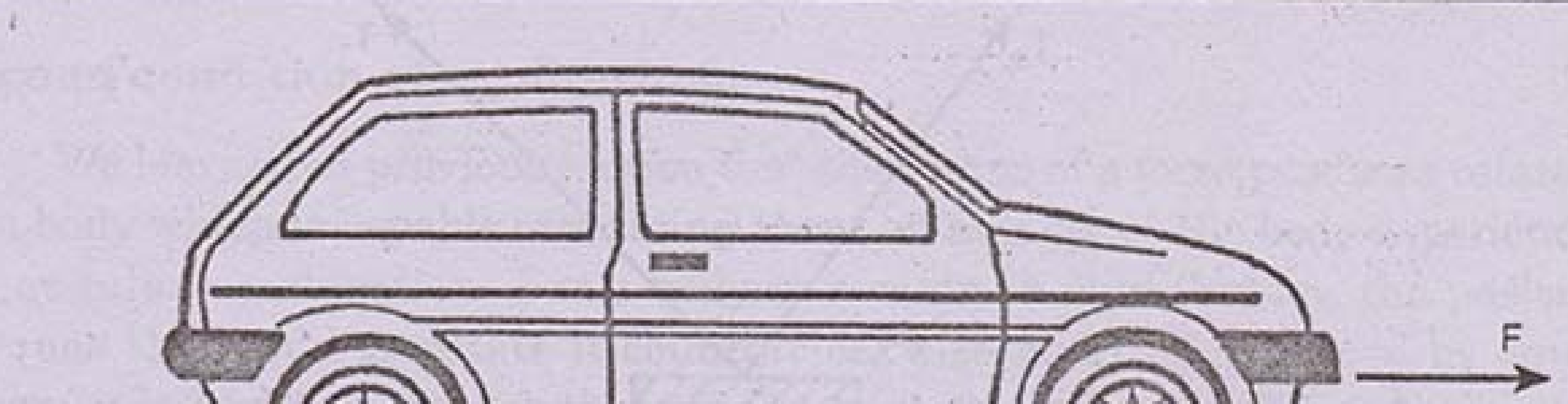


Fig. 6.13

6.7 CONDITIONS OF EQUILIBRIUM:

For a body to remain in equilibrium, it must satisfy two conditions, known as (i) First condition of equilibrium (ii) Second condition of equilibrium.

First condition of equilibrium:

The first condition of equilibrium states that if a body is in equilibrium, then the vector sum of all the forces acting on the body must be equal to zero. This can be written as

$$\Sigma\vec{F} = \vec{F}_1 + \vec{F}_2 + \vec{F}_3 + \text{-------} = 0 \text{ --------- } (6.1)$$

The Greek letter sigma, Σ, means summation. Thus $\Sigma\vec{F}$ represents the summation of all the forces $\vec{F}_1$, $\vec{F}_2$, $\vec{F}_3$, ------- acting on the body. It follows from the above statement that the linear acceleration of a body in equilibrium is zero. In two dimensions the first condition of equilibrium leads to the following

$$\Sigma\vec{F}_x = \vec{F}_{1x} + \vec{F}_{2x} + \vec{F}_{3x} + \text{-------} = 0 \text{ --------- } (6.2a)$$

$$\Sigma\vec{F}_y = \vec{F}_{1y} + \vec{F}_{2y} + \vec{F}_{3y} + \text{-------} = 0 \text{ --------- } (6.2b)$$

i.e. if a body is in equilibrium, then the summation of the x-components ($\vec{F}_{1x}$, $\vec{F}_{2x}$, $\vec{F}_{3x}$, ------) of all the forces acting on the body must be equal to zero, and the summation of the y-components ($\vec{F}_{1y}$, $\vec{F}_{2y}$, $\vec{F}_{3y}$, ------) of the forces must be equal to zero.

In physics the first condition of equilibrium is used to determine the magnitude and direction of an unknown force acting on a body if all the other forces are known.

Example 6.2: A block of mass 100 kg is suspended as shown in Fig. 6.14. Find the tensions T_1 and T_2 in the two ropes. Take the value of $g = 9.8\ m/s^2$.

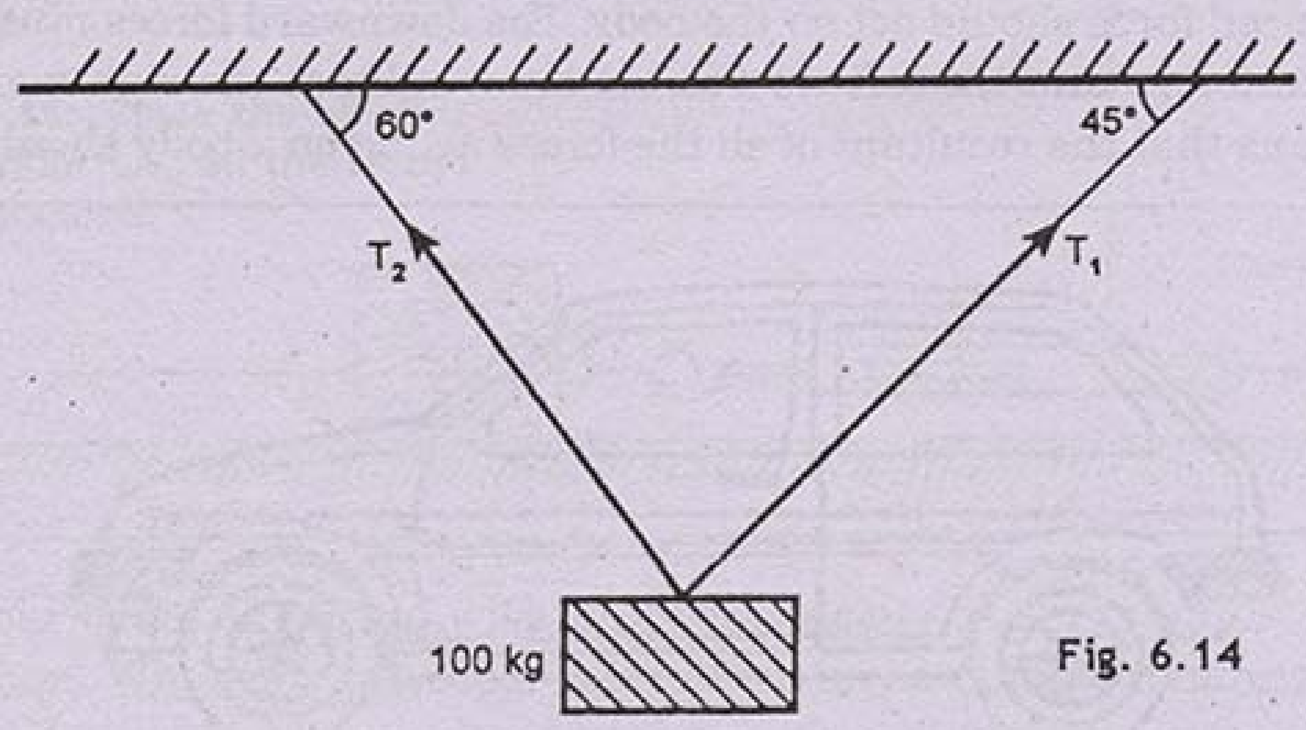
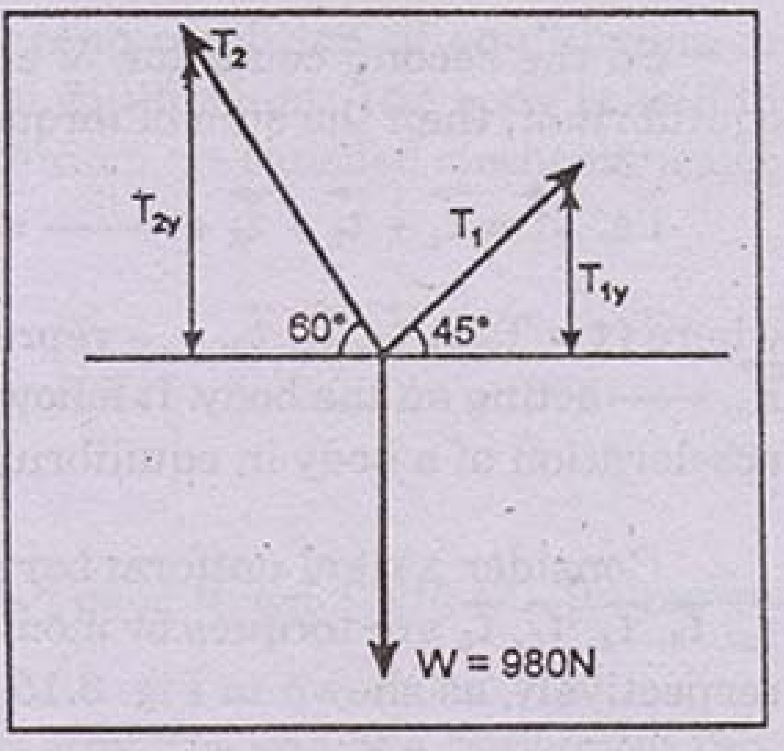


Fig. 6.14

Solution:

Resolve the forces (tensions) T_1 and T_2 into the rectangular components.



$$T_{1x} = T_1 \cos 45° \quad , \quad T_{1y} = T_1 \sin 45°$$

$$T_{2x} = T_2 \cos 60° \quad , \quad T_{2y} = T_2 \sin 60°$$

Applying the first condition of equilibrium, we get

$$\Sigma F_x = 0$$

$$T_{1x} - T_{2x} = 0$$

or $\quad T_1 \cos 45° - T_2 \cos 60° = 0$

or $\quad 0.707\, T_1 \quad - 0.5\, T_2 \qquad = 0$ ---------------(1)

$$\Sigma F_y = 0$$

$$T_{1y} + T_{2y} - w = 0$$

or $\quad T_1 \sin 45° + T_2 \sin 60° - 980 = 0$

or $\quad 0.707\, T_1 \quad + 0.866\, T_2 = 980$ ------------(2)

Substituting the value of T_1 from eq.(1) to eq.(2), we get

$$0.5\, T_2 + 0.866\, T_2 = 980$$

or $\quad 1.366\, T_2 = 980$

or $\quad T_2 = 717.4$ N

Substituting the value of T_2 in eq.(1), we get

$$0.707\, T_1 - 0.5 \times 717.4 = 0$$

or $\quad T_2 = 507.4$ N

Second condition of equilibrium:

We learned in previous section that the torque of a force produces rotation in a body which is capable of rotating about an axis. Thus the body experiences an angular acceleration. For a body to remain in equilibrium, the positive torques that tend to rotate it counterclockwise must be balanced by equal magnitude negative torques that tend to rotate it clockwise. This is the second condition of equilibrium.

So the second condition of equilibrium can be stated as, if a body is in equilibrium, then the sum of torques acting on the body must be equal to zero.

i.e. $\Sigma\vec{\tau} = \vec{\tau}_1 + \vec{\tau}_2 + \vec{\tau}_3 + \text{--------} = 0$ ------- (6.3)

where (τ = Tau) $\vec{\tau}_1, \vec{\tau}_2, \vec{\tau}_3$, ----- represent the torques caused by all forces $\vec{F}_1, \vec{F}_2, \vec{F}_3$, ------ acting on the body. It follows from the above statement that the angular acceleration of a body in equilibrium is zero.

Consider a rigid uniform bar supported at its middle point O. Suppose $\vec{\tau}_1, \vec{\tau}_2, \vec{\tau}_3, \vec{\tau}_4, \vec{\tau}_5, \vec{\tau}_6$ are torques or moments produced by forces $\vec{F}_1, \vec{F}_2, \vec{F}_3, \vec{F}_4, \vec{F}_5, \vec{F}_6$ respectively, as shown in Fig. 6.15.

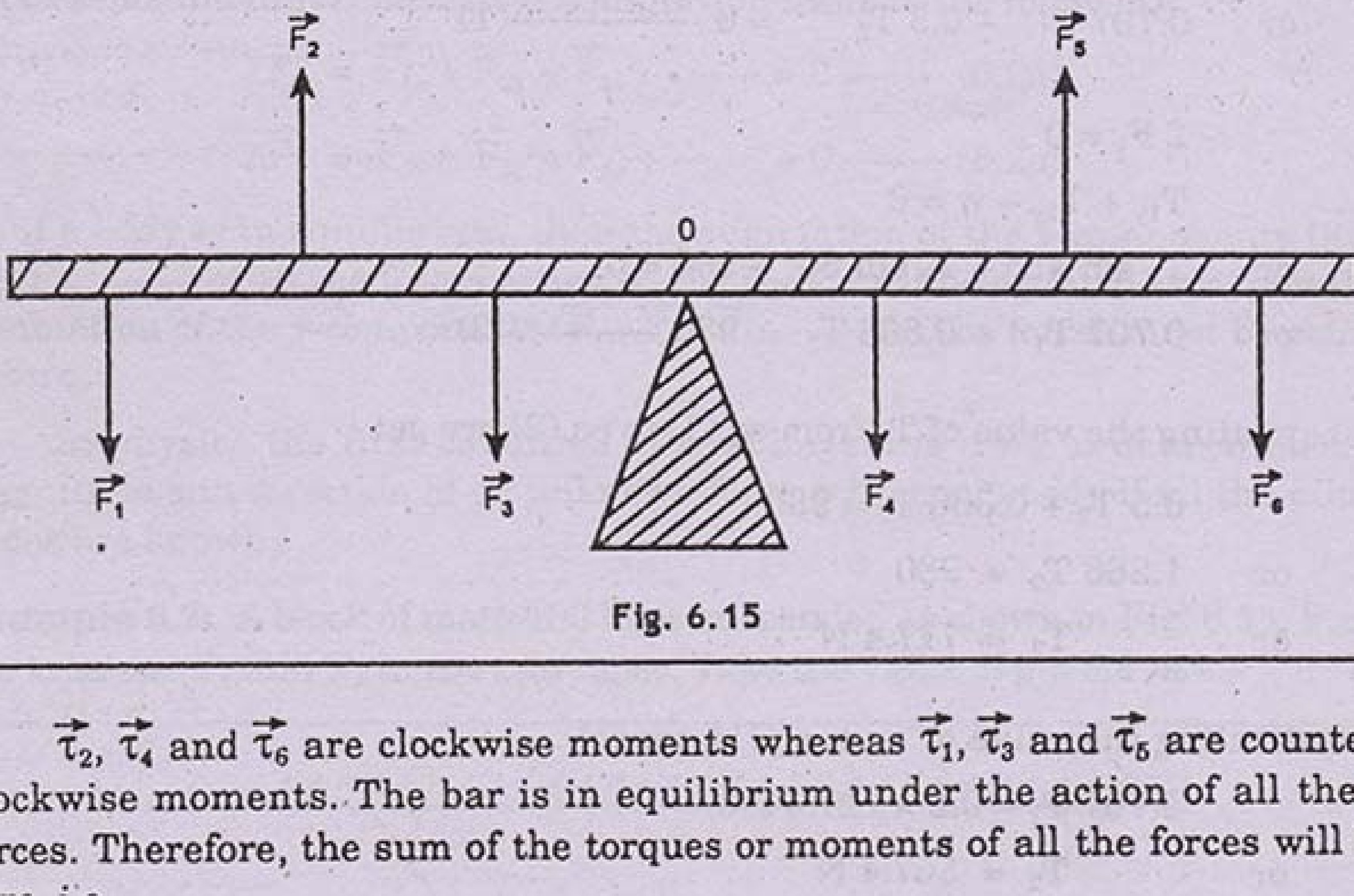


Fig. 6.15

$\vec{\tau}_2, \vec{\tau}_4$ and $\vec{\tau}_6$ are clockwise moments whereas $\vec{\tau}_1, \vec{\tau}_3$ and $\vec{\tau}_5$ are counter-clockwise moments. The bar is in equilibrium under the action of all these forces. Therefore, the sum of the torques or moments of all the forces will be zero, i.e.

$$\vec{\tau}_1 + (-\vec{\tau}_2) + \vec{\tau}_3 + (-\vec{\tau}_4) + \vec{\tau}_5 + (-\vec{\tau}_6) = 0$$

$$\therefore\ \vec{\tau}_1 - \vec{\tau}_2 + \vec{\tau}_3 - \vec{\tau}_4 + \vec{\tau}_5 - \vec{\tau}_6 = 0$$

$$\therefore\ \vec{\tau}_1 + \vec{\tau}_3 + \vec{\tau}_5 = \vec{\tau}_2 + \vec{\tau}_4 + \vec{\tau}_6$$

$\therefore$ Sum of the counter-clockwise moments

= Sum of the clockwise moments

This verifies the principle of moments.

When the first condition of equlibrium is satisfied, the body is said to be in translational equilibrium. When the second condition of equilibrium is satisfied, the body is said to be in rotational equilibrium. The body is said to be in equilibrium if the two conditions of equilibrium are satisfied mathematically, it can be written as

$$\left.\begin{aligned}\Sigma F_x &= 0\\ \Sigma F_y &= 0\end{aligned}\right\} \text{ First condition}$$

$$\Sigma\tau = 0 \quad \text{Second condition}$$

Second condition of equilibrium is also used to calculate the unknown force acting on bodies in equilibrium.

Example 6.3: A lamp weighing 10N hangs from the end of a horizontal rod, 10m long, sticking out perpendicular from a wall. The other end of the rod is hinged to the wall. A wire is attached to the middle of the rod and to a hook in the wall 5m above the hinge. Find the tension in the wire if the rod weighs 5N.

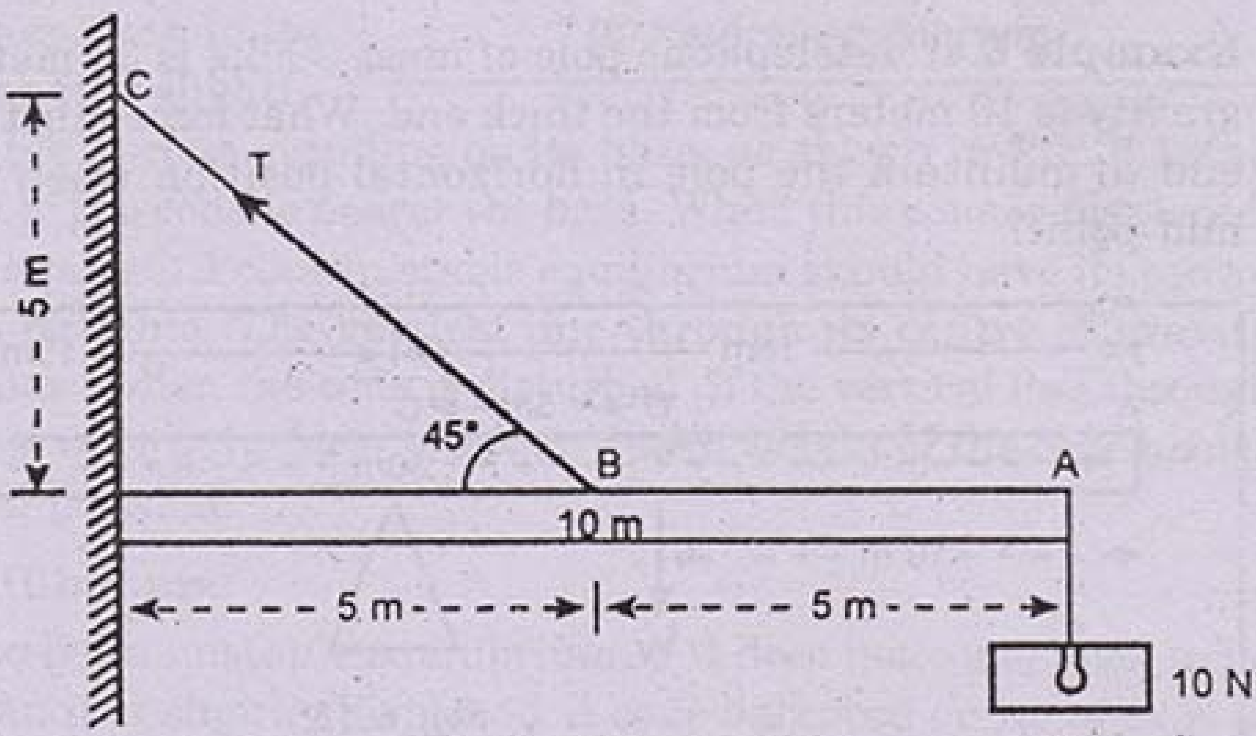


Fig. 6.16

Solution:

Suppose the rod OA is hinged to the wall at point 0. The lamp is hanged at point A, B is mid-point of the rod where the weight of the rod acts. C is

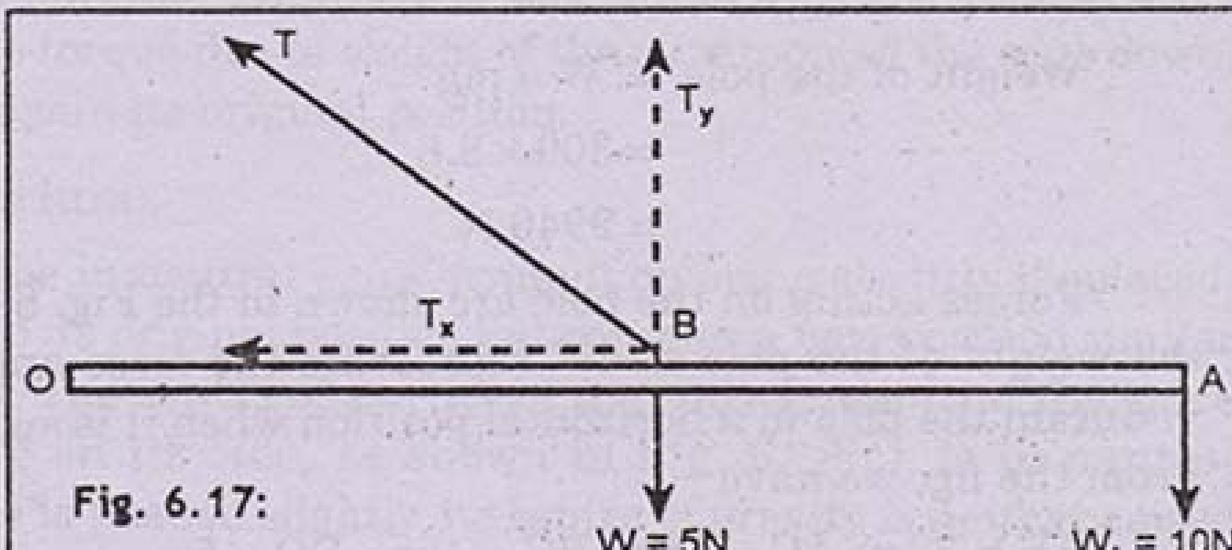


Fig. 6.17:

the point on the wall which is hooked by a wire to the point B. Forces acting on the rod are shown in the fig.

From the Fig. 6.16, we get

$OA = 10m, AB = 5m, OB = 5m, OC = 5m$

$\therefore \angle OBC = 45°$

By resolving tension T in the wire into rectangular components, we get

$T_x = T \cos 45° \quad , \quad T_y = T \sin 45°$

By applying second condition of equilibrium about 0, we get

$$\Sigma\tau = 0$$

$$T_y \times OB - W \times OB - W_1 \times OA = 0$$

or $T \sin 45° \times 5 - 5 \times 5 - 10 \times 10 = 0$

or $0.707 \times 5T - 25 - 100 = 0$

or $3.535T - 125 = 0$

or $T = 35.36\ N$

Example 6.4: A telephone pole of mass 300kg is 30 meters long. Its centre of gravity is 10 meters from the thick end. What force must be applied at the thin end to maintain the pole in horizontal position when it is supported at its mid-point?

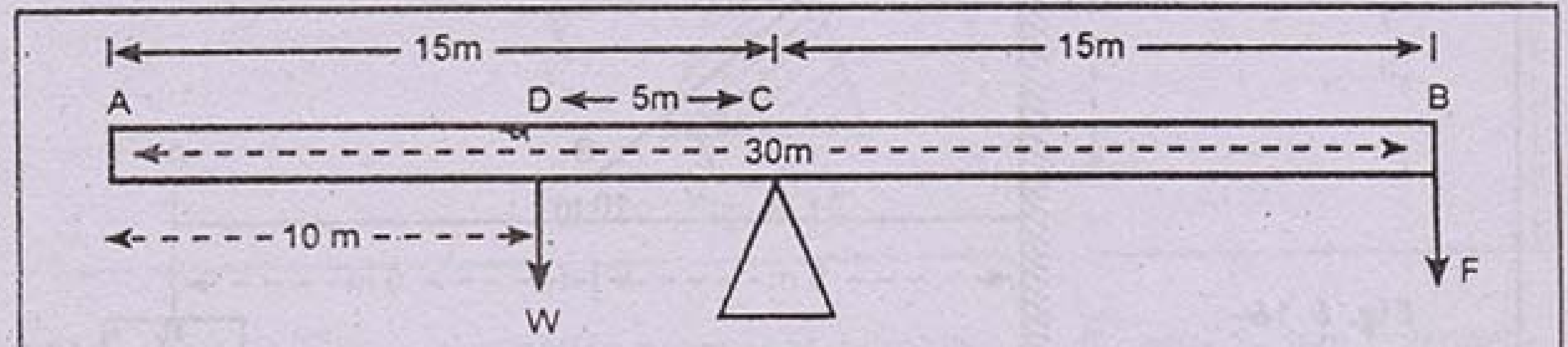


Fig. 6.18

Solution:

Weight of the pole $= W = mg$

$= 300 \times 9.8$

$= 2940\ N$

Forces acting on the pole are shown in the Fig. 6.18, where W represents the weight of the pole and F represents the force applied at the thin end to maintain the pole in a horizontal position when it is supported at its mid-point. From the fig, we have

$AB = 30m, AC = 15m, BC = 15m, DC = 5m$

By applying second condition of equilibrium about C, we get;

$$\Sigma\tau = 0$$

$$W \times DC - F \times BC = 0$$

or $$2940 \times 5 - F \times 15 = 0$$

or $$15F = 2940 \times 5$$

or $$F = 980\text{ N in the downward direction}$$

6.8 STATES OF EQUILIBRIUM:

There are three states of equilibrium of a body. They are (i) stable equilibrium, (ii) unstable equilibrium, (iii) neutral equilibrium.

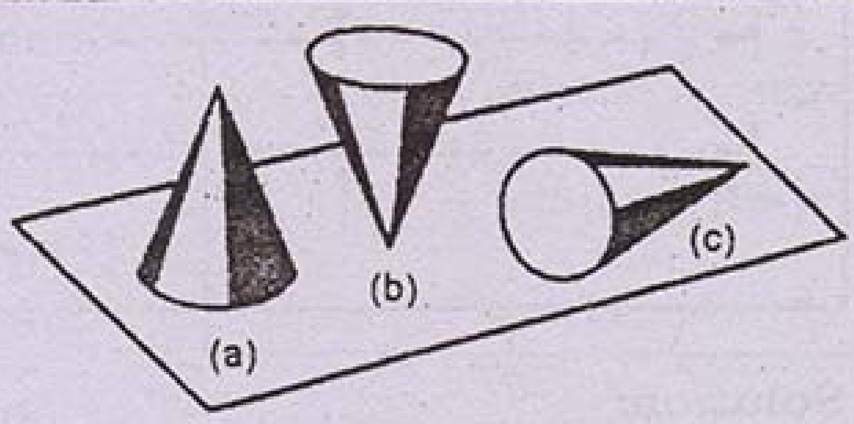


Fig. 6.19: (a) stable; (b) unstable (c) neutral equilibrium

(i) Stable equilibrium:

A body is said to be in stable equilibrium if it comes back to its original position when it is slightly displaced. For example, a cone standing on its base, as shown in Fig. 6.19(a). The centre of gravity of the cone is nearer the base. When this cone is displaced, its centre of gravity is raised. A cone in stable equilibrium should have its centre of gravity as low as possible. The vertical line through its centre of gravity should be within its base when the cone is disturbed. If the vertical line through the centre of gravity is within the base, a torque due to weight of the cone brings back it to the original position.

(ii) Unstable equilibrium:

A body is said to be in unstable equilibrium, if it does not come back to its original position when it is slightly displaced. A cone balanced on its apex is in unstable equilibrium, as shown in Fig. 6.19(b). When this come is disturbed, its centre of gravity is lowered. The vertical line through its centre of gravity is outside the base. The torque due to weight of the cone toppled the cone down. So the cone does not regain its original position.

(iii) Neutral equilibrium:

A body is said to be in neutral equilibrium if on being slightly displaced, it does not come back to its original position but occupies a new position similar to its original position. The centre of gravity of the body remains at the same height. A cone resting on its side, as shown in Fig. 6.19(c), is in nautral equilibrium. If the cone is pushed slightly, its centre of gravity is neither raised nor lowered but it remains at the same height.

Example 6.5: A uniform rod 10 meters long and weighing 30N is supported in a horizontal position on a fulcrum with weights of 40N and 50N suspended from its ends, as shown in Fig. 6.20. Find the position of the fulcrum.

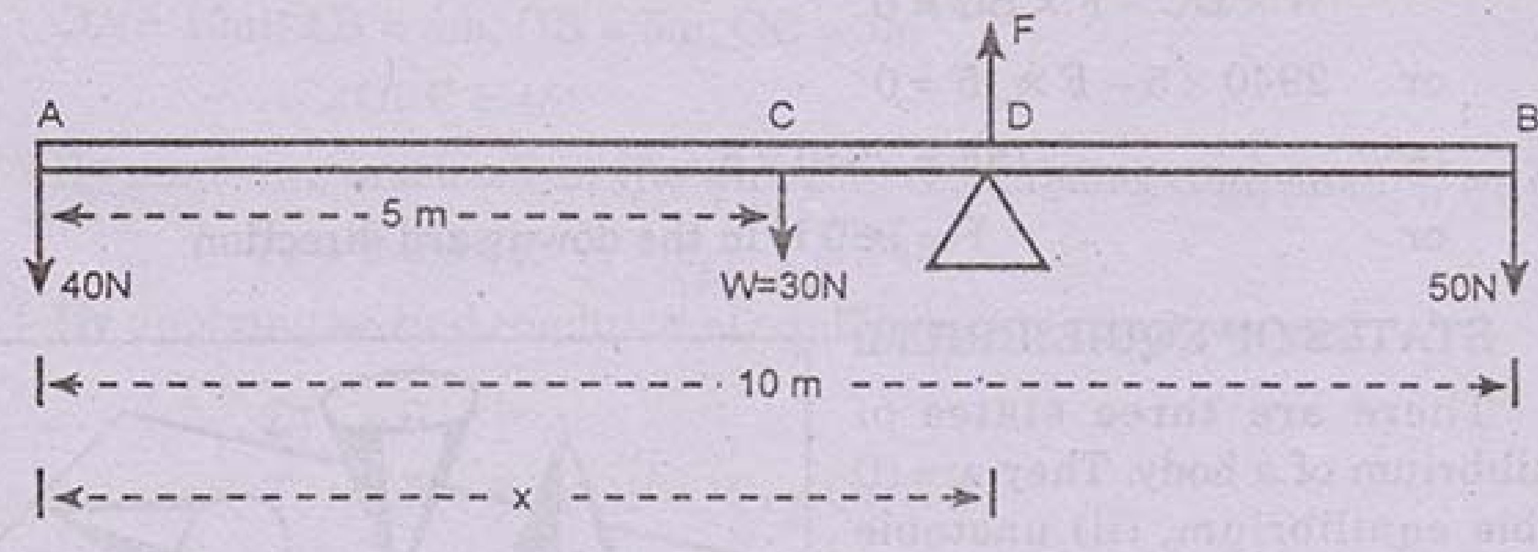


Fig. 6.20

Solution:

The weight-W of the uniform rod acts at C. Let D be the fulcrum and F be the upward force at the fulcrum. Suppose x is the position of fulcrum with respect to A. The rod is in equilibrium, apply first condition of equilibrium, we get

$\Sigma F_x = 0$, No force along the x-axis

$\Sigma F_y = 0$

$F - 40 - 30 - 50 = 0$

or $F = 120\text{ N}$

By applying second condition of equilibrium about A, we get

$\Sigma \tau = 0$

$F \times AD - W \times AC - 50 \times AB = 0$

or $120 \times x - 30 \times 5 - 50 \times 10 = 0$

or $120\,x - 150 - 500 = 0$

or $120\,x = 650$

or $x = 5.4\text{m}$

Example 6.6: A uniform beam 15m long weighs 100N and is arranged as shown in Fig. 6.21. Find (a) the tension in the string and (b) the horizontal and vertical components of the force exerted on the beam at the hinge.

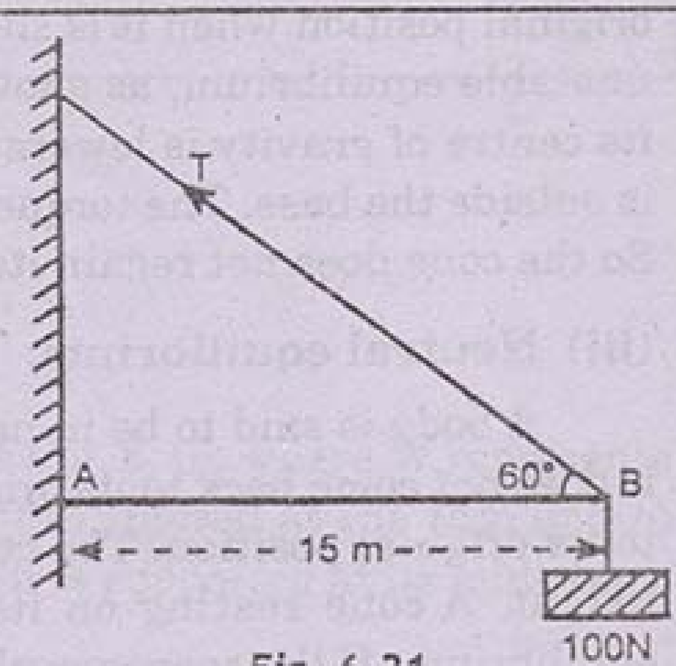


Fig. 6.21

Solution:

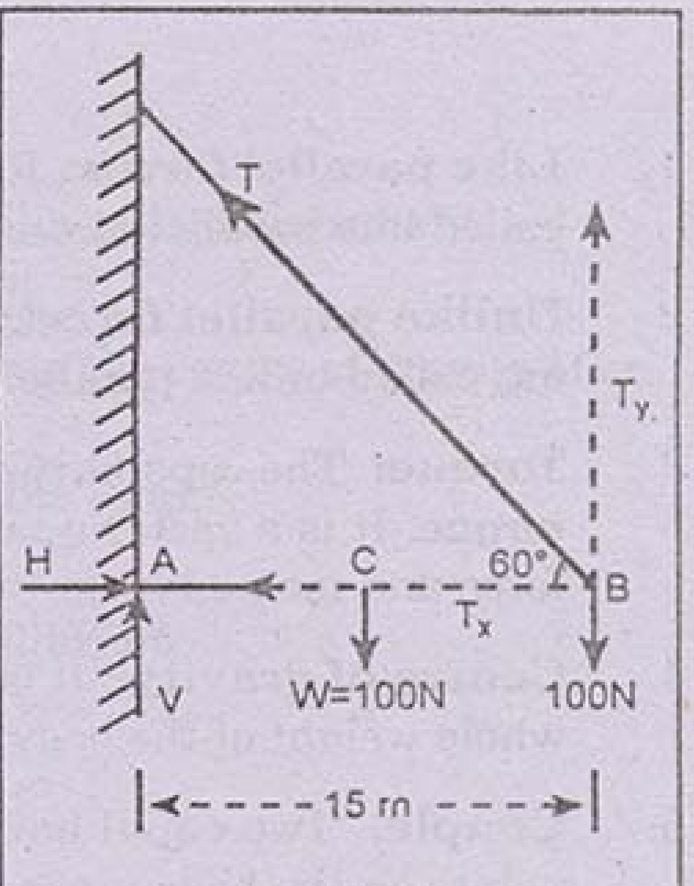


Fig. 6.22

Suppose H and V are the horizontal and vertical components of the force exerted on the beam at the hinge. Let T be tension in the string. Resolve T into the rectangular components.

$$T_x = T\cos 60^\circ \quad , \quad T_y = T\sin 60^\circ$$

By applying first condition of equilibrium, we get

$$\Sigma F_x = 0$$

$$H - T\cos 60^\circ = 0$$

or $\quad H = T \times 0.5$

or $\quad H = 0.5\,T$ ------------------------- (1)

$$\Sigma F_y = 0$$

$$V + T\sin 60^\circ - 100 - W = 0$$

or $\quad V + T \times 0.866 - 100 - 100 = 0$

or $\quad V + 0.866\,T - 200 = 0$

or $\quad V = 200 - 0.866\,T$ -------------- (2)

By applying second condition of equilibrium about A, we get

$$\Sigma \tau = 0$$

$$T\sin 60^\circ \times AB - 100 \times AB - W \times AC = 0$$

or $\quad T \times 0.866 \times 15 - 100 \times 15 - 100 \times 7.5 = 0$

or $\quad 12.99\,T - 1500 - 750 = 0$

or $\quad 12.99\,T = 2250$

or $\quad T = 173.2\text{ N}$

By substituting the value of T in eq.(1), we get

$$H = 0.5 \times 173.2$$

or $\quad H = 86.6\text{ N}$

By substituting the value of T in eq.(2), we get

$$V = 200 - 0.866 \times 173.2$$

or $\quad V = 200 - 150$

or $\quad V = 50\text{ N}$

SUMMARY

1. **Like parallel forces:** Parallel forces which act in the same direction are called like parallel forces.

2. **Unlike parallel forces:** Parallel forces which act in opposite direction are called unlike parallel forces.

3. **Torque:** The capability of a force to rotate a body about a point is called torque. It is a vector quantity and its magnitude equal to the product of force and moment arm.

4. **Centre of gravity:** It is a point inside or outside the body at which the whole weight of the body is acting.

5. **Couple:** Two equal but opposite parallel forces which act at different points on the body make together a couple.

6. **Equilibrium:** A body at rest or moving with uniform velocity is in a state of equilibrium. The net force on such a body is zero and its acceleration is also zero.

7. **Conditions of equilibrium:** For a body to be in a state of equilibrium two conditions are to be fulfilled.

   (i) The resultant of all the forces acting on the body should be zero.

   (ii) The resultant of all the torque acting on the body should be zero.

8. **States of equilibrium:**

   (i) **Stable equilibrium:** Such a state in which on disturbing a body slightly, the body comes back to its original position. The body is said to be in stable equilibrium.

   (ii) **Unstable equilibrium:** Such a state in which on disturbing the body slightly, the body does not come to its original position. The body is said to be in unstable equilibrium.

   (iii) **Neutral equilibrium:** A state in which on disturbing the body slightly, the centre of gravity does not shift. The body is said to be in the state of neutral equilibrium.

QUESTIONS:

6.1 Answer the following questions:

1. What do you understand by two like and unlike parallel forces?
2. Discuss torque or moment of a force in detail.
3. Define centre of gravity. How would you locate the centre of gravity of an irregular piece of a metal sheet?
4. What is a couple? Calculate the moment of the couple.
5. Define static and dynamic equilibrium.
6. State and explain the two conditions of equilibrium.
7. Explain three states of equilibrium.

6.2 Fill in the blanks.

1. If two parallel forces have opposite directions, they are called ________ parallel forces.
2. An axis, about which a body is free to rotate, is called ________.
3. Torque about an axis is the product of the ________ and ________.
4. Torque is a ________ quantity.
5. Conventionally the torque, producing counter-clockwise rotation is taken as ________.
6. In S.I. units, the unit of torque is ________.
7. If the net force on a body is zero, it is said to be in ________ equilibrium.
8. If the net torque on a body is zero, it is said to be in ________ equilibrium.
9. Two equal, parallel, and unlike forces having different lines of action, is called a ________.
10. ________ is equal to the product of one of the forces and the perpendicular distance between the two forces.
11. A couple can be balanced by an ________ and ________ couple.
12. ________ is that point on which the weight of the body acts.
13. The centre of gravity of a sphere lies at its ________.
14. A body at rest is said to be in ________.
15. A body in uniform motion along a straight line is said to be in ________ equilibrium.
16. There are ________ states of equilibrium.
17. If a body in unstable equilibrium is slightly disturbed, it does not come back to its ________ position.

6.3 Tick (✓) the correct answer.

1. If two parallel forces have the same directions, they are called
   (a) unlike parallel forces (b) like parallel forces
   (c) rectangular forces (d) couple.

2. The turning effect of a force about an axis is
   (a) force (b) rotation
   (c) torque (d) momentum

3. Clockwise torque is considered as
   (a) positive torque (b) negative torque
   (c) unit torque (d) zero torque

4. Torque is a __________.
   (a) scalar quantity (b) vector quantity
   (c) negative quantity (d) none of these

5. The centre of gravity of a uniform rod is
   (a) end of rod (b) centre of rod
   (c) both (a) & (b) (d) none of these

6. A body at rest or moves with uniform velocity is said to be
   (a) motion (b) equilibrium
   (c) static equilibrium (d) dynamic equilibrium

7. If vector sum of all the forces on a body is equal to zero, then it is said to be in
   (a) translational equilibrium (b) rotational equilibrium
   (c) static equilibrium (d) dynamic equilibrium

8. A body will be in rotational equilibrium if the sum of clockwise torques acting on the body is equal to the sum of _________.
   (a) perpendicular forces (b) parallel forces
   (c) counter clockwise torques (d) clockwise torques

9. The centre of gravity of a body is a point where ________ acts.
   (a) the torque (b) the external force
   (c) the weight of the body (d) none of these

10. The first condition of equilibrium states that
    (a) $\Sigma P = 0$ (b) $\Sigma \tau = 0$
    (c) $\Sigma F = 0$ (d) both (b) & (c)

11. Two forces which are equal and opposite in direction and not acting along the same line constitute a
    (a) torque (b) couple
    (c) resultant vector (d) same vector

12. If a body comes back to its original position when it is slightly displaced, it is said to be in
    (a) stable equilibrium (b) unstable equilibrium
    (c) neutral equilibrium (d) none of these

13. A cone resting on its side is in ________ equilibrium.
    (a) neutral (b) stable
    (c) unstable (d) dynamic

14. The unit of torque in S.I. units is
    (a) Newton (b) Meter
    (c) Newton-meter (d) Pound

6.4 Write "True" or "False" against each statement.

1. Two unlike parallel forces have same directions.
2. The centre of gravity of a hollow sphere is not within the material.
3. Torque is a scalar quantity.
4. Torque depends upon the magnitude of force and moment arm.
5. The first condition of equilibrium states that the sum of torques acting on a body is equal to zero.
6. A body at rest is said to be in dynamic equilibrium.
7. The second condition of equilibrium states that the sum of counter-clockwise torques is equal to the sum of clockwise torques.
8. In stable equilibrium the body does not come back to its original position when it is slightly displaced.
9. A body is said to be in translational equilibrium if its linear acceleration is equal to zero.
10. A couple can not be balanced by a single force.

PROBLEMS

6.1 A force of 25N acts on a body. If moment arm is 2m, find the value of torque.

(50 N-m)

6.2 A force is applied perpendicularly on a door, 4 metres wide which requires a torque of 120 N-m to open it. What will be minimum force required?

(30 N)

6.3 What is the moment of the couple of 10N acting at the extremities of a rod 5m long, as shown in Fig. P 6.1. How can this couple be balanced?

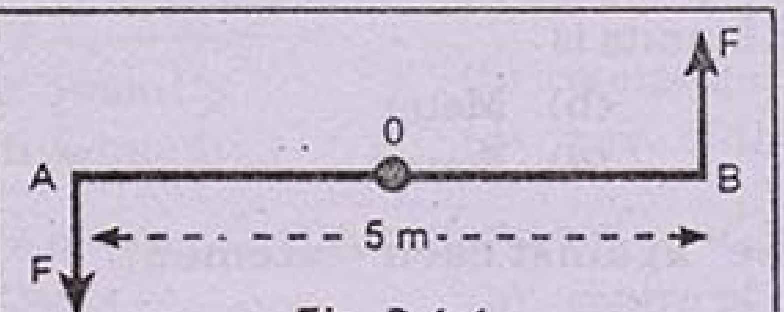
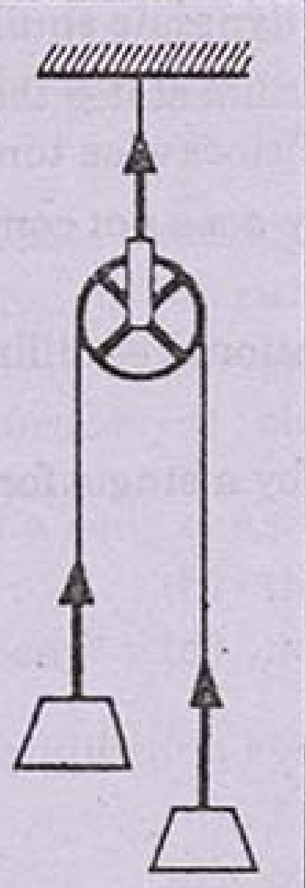


Fig. P-6.1

(50 N-m counter-clockwise)

6.4 Two 10Kg masses are suspended at the two ends of a rope which passes over a light frictionless pulley. The pulley is attached to a chain which goes to the ceiling, as shown in Fig. P 6.2. What is the tension in the (a) rope (b) chain?

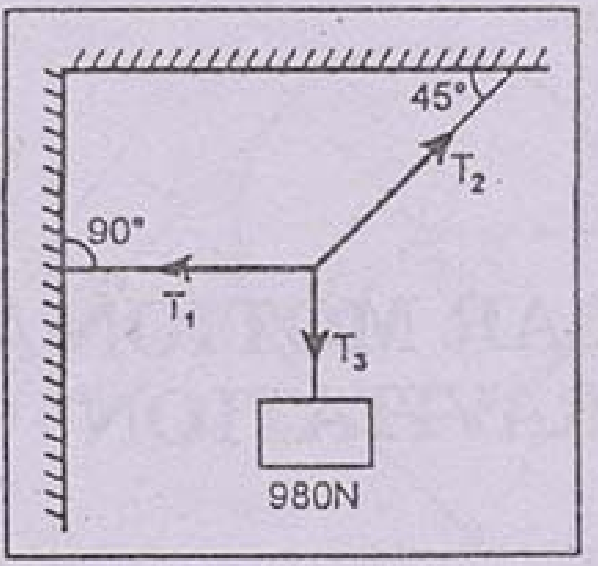
Fig. P-6.2



[(a) 98N, (b) 196N]

6.5 Find the tension in each cord in Fig. P 6.3 if the weight of the suspended body is 980 N.

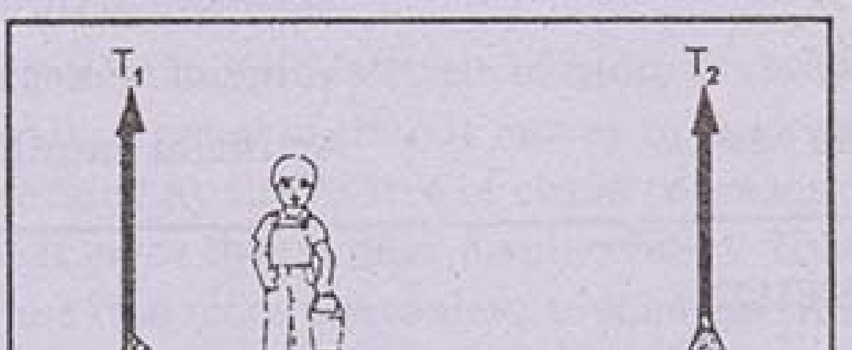
Fig. P-6.3



(T_1=980N, T_2=1386N, T_3=980N)

6.6 A painter weighing 150N is standing on a uniform plank 10m long at a distance 2m from one end of the plank. The weight of the plank is 50N and it is supported by two ropes at the ends, as shown in Fig. P 6.4. Find the tension in the ropes.

Fig. P-6.4



(T_1=55N, T_2=145N)

6.7 Find the tension in the cable and the components of the force exerted on the beam by the wall in Fig. P 6.5. The weight of the beam is 300N.

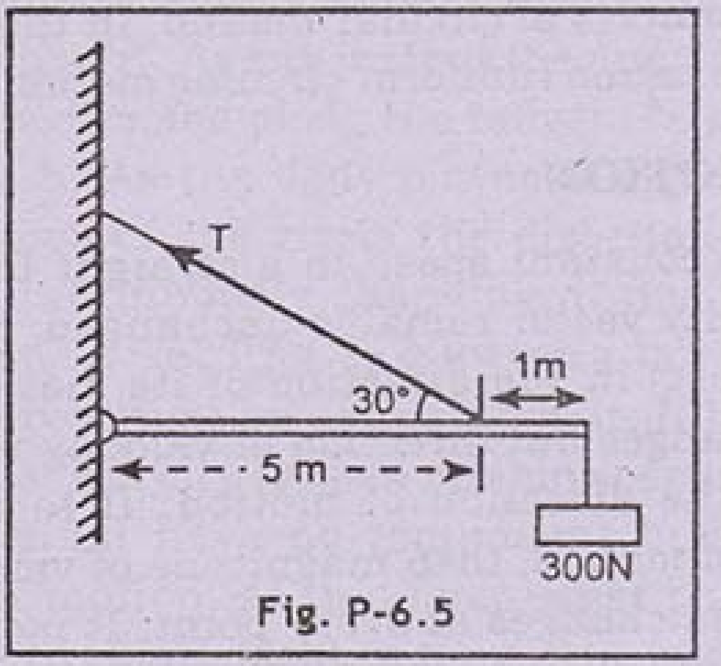


Fig. P-6.5

(T=1080N, H=935.28N, V=60N)

CHAPTER – 7

CIRCULAR MOTION AND GRAVITATION

LEARNING OBJECTIVES:

- Introduction.
- Uniform circular motion.
- Centripetal acceleration.
- Centripetal force.
- Centrifugal force.
- Banking of the roads.
- Centrifuge.
- Law of universal gravitation.
- Gravitational constant "G".
- Mass of earth.
- Value of 'g' decreases with altitude.
- Artificial satellites.

7.1 INTRODUCTION

It has been said that nature is in unending motion. In previous chapter we have discussed the motion of body in straight line. To study various physical phenomenon around us, we certainly would take to describe circular motion.

In our daily life we see the objects in rotation. The rotation of our earth and other planets around the sun, the rotation of moon around earth, the wheel of moving bicycle are common examples of circular motion. In this chapter we shall study the simplest form of rotation (uniform circular motion).

7.2 UNIFORM CIRCULAR MOTION

When an object moves at a constant speed in a straight line, both the magnitude and direction of velocity vector remains unchanged, however, if a force is acting on the body (object) in the direction of its motion than the magnitude of velocity (speed) changes but direction of velocity remain same and acceleration produced in the direction of motion. If force is acting perpendicularly to the direction of motion, then magnitude of velocity (speed) remains the same but its direction changes at every point. It means there is change in velocity. There must be some acceleration and object due to this force moves in curve path. The simplest type of curve path is circle.

"When a body moves in a circular path with constant speed it is said to describe uniform circular motion".

At any point on the circle the direction of velocity is directed along the tangent to the circle at that point. If at any point a body is made free of force it will move tangential to the circle at that point. But as the body moving along the circumference of the circle there must be a force acting towards the centre of the circle and this force is called centripetal force.

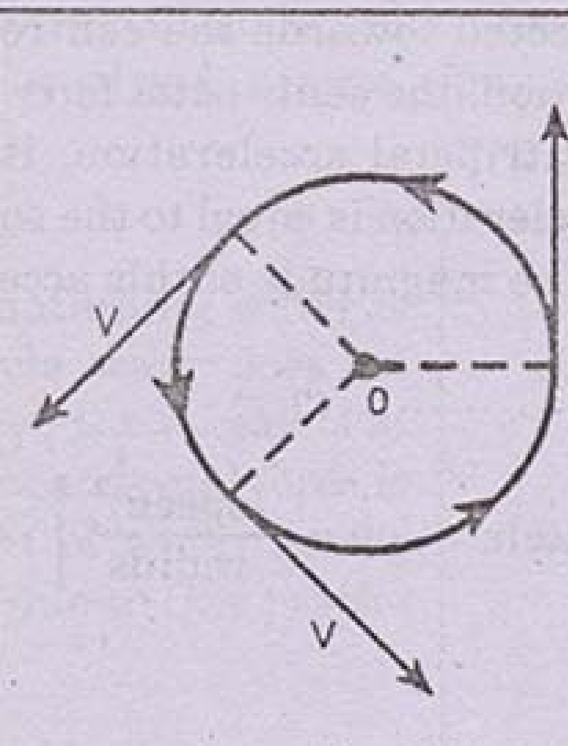


Fig. 7.1

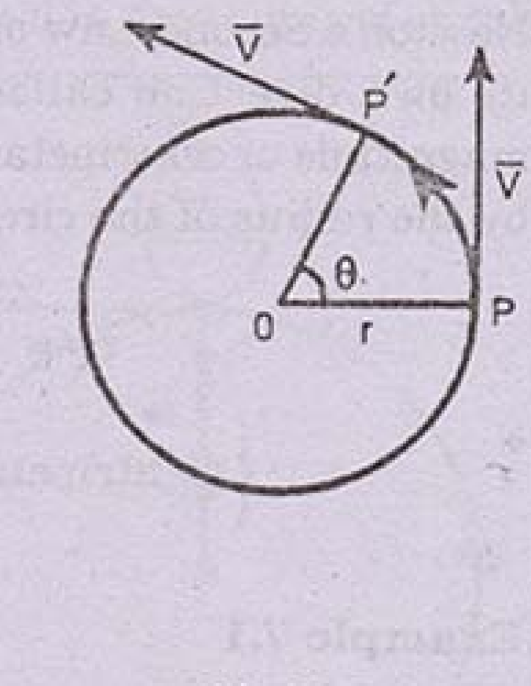


Fig. 7.2

Consider a body revolving in a circle of radius "r". Let the body be at point P initially. After small interval of time it moves to position P′. The angle ∠POP′ or 'θ' (Fig. 7.2) subtended at the centre of circle represents the turning of body during this interval. It gives the angular displacement. The angular displacement measure in degree but it is more convenient to measure it in an other unit called "radian". One radian is the angle subtended at the centre of a circle by an arc equal in length to the radius of the circle.

1 radian = 57.3°

7.3 CENTRIPETAL ACCELERATION

Consider a body of mass "m" moving with constant speed "v" around a circle of radius "r". At any instant the direction of velocity is directed along the tangent to the circular path. As the body moves along the circumference of the circle the direction of velocity is continuously changing, it has an acceleration. According to Newton's law of motion a force must be acting upon it. If the magnitude of velocity (speed) remain constant this force must have no component in the direction of motion, so this force must be at right angle to the tangent to the circle, and it is directed towards the centre of the circle called

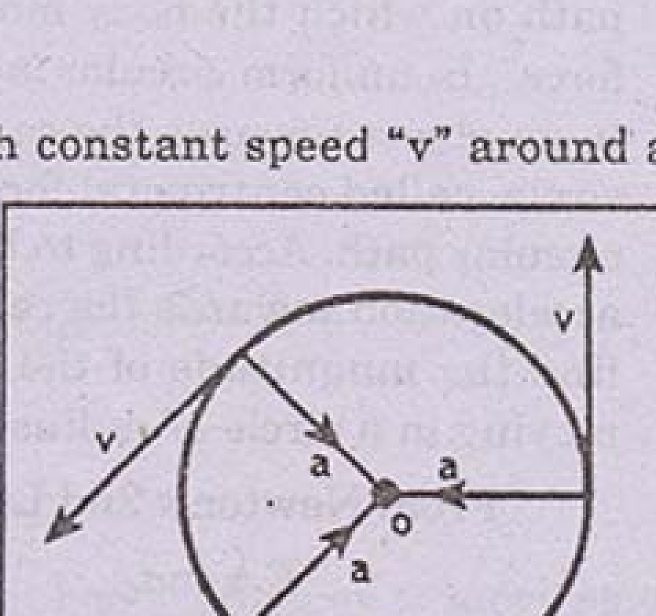


Fig. 7.3

centripetal force (force directed towards the centre of circle). According to Newton's Second Law of Motion, the centripetal force produces acceleration in its own direction called centripetal acceleration. It can be shown that the magnitude of centripetal acceleration is equal to the square of the speed divided by the radius of the circle. The magnitude of this acceleration "a_c" is given by

$$a_c = \frac{v^2}{r} \text{ .... } (7.1)$$

$$\left\{ \text{Centripetal acceleration} = \frac{(\text{speed})^2}{\text{radius}} \right\}$$

Example 7.1

The moon revolves around the earth in an orbit that is approximately circular and of mean radius 3.84×10^8m. The speed of moon is 1022 m/s. What is the centripetal acceleration of the moon.

Solution:

To calculate the centripetal acceleration

$$a_c = \frac{v^2}{r}$$

$$a_c = \frac{(1022)^2 \text{m/s}}{3.84 \times 10^8 \text{m}} = 2.72 \times 10^{-3} \text{m/s}^2$$

7.4 CENTRIPETAL FORCE:

"The force which acts towards the centre along the radius of a circular path on which the body moving with a uniform velocity is called centripetal force". In uniform circular motion there must be a net force that must be applied to an object towards the centre of a circle to make the object move round the circle, called centripetal force 'w', without this force object does not move in a circular path. According to Newton's Second Law of Motion, this force produces acceleration towards the centre of the circle called centripetal acceleration. To find the magnitude of the centripetal force, consider an object of mass "m" moving in a circle of radius "r" with constant speed "v".

From Newton's 2nd Law of Motion, we have

$$F = ma_c$$

Where "a_c" is the centripetal acceleration produced by centripetal force "F". Substituting the value of "a_c" in above equation, we have centripetal force

$$F = \frac{mv^2}{r} \text{ .... } (7.2) \quad \because a_c = \frac{v^2}{r}$$

Thus the centripetal force is directly proportional to the mass of the object, to the square of its linear speed and inversally proportional to the radius of circular path.

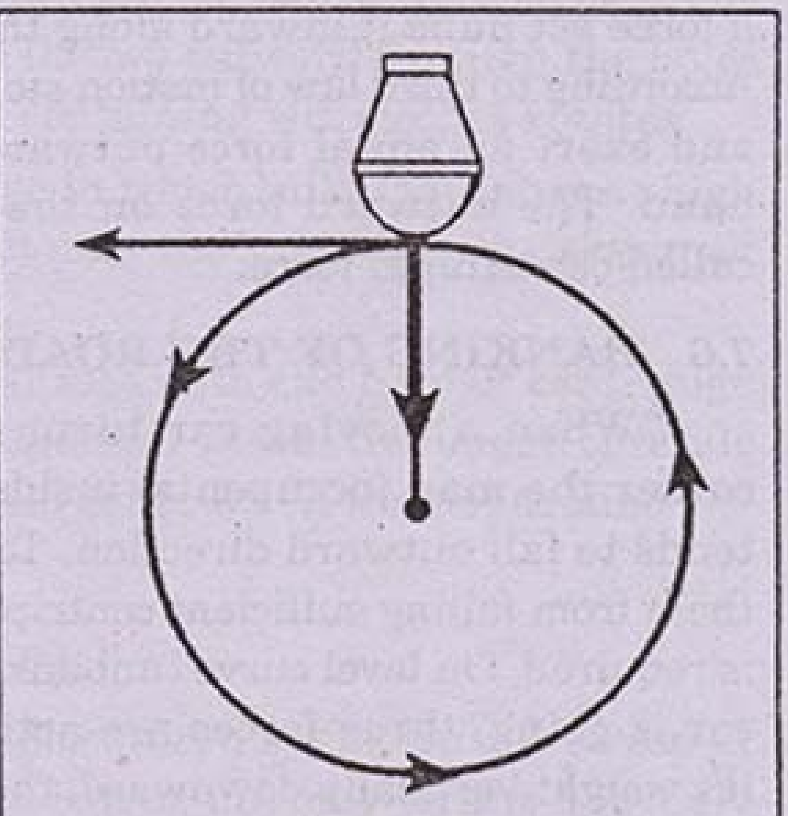


Fig. 7.4

Example 7.2

When a bucket containing water is rotated in a vertical circle, water does not fall downword even when the bucket is at the highest point. The centripetal force is provided by the weight of water acting vertically downwords, due to this reason water does not fall.

Example 7.3

In the case of motion of the moon round the earth centripetal force is provided by the gravitational force of attraction of the earth on the moon. Similarly the force of attraction of sun on the planets provide the necessary centripetal force and the planets revolve round the sun.

Example 7.4

A car of mass 1500 Kg moving in a circular track of radius 20m at a constant speed 20 m/sec. Find the centripetal force required for this purpose.

Solution:

$$m = 1500\ \text{Kg}$$
$$v = 20\ \text{m/s}$$
$$r = 20\ \text{m}$$

$$F_c = \frac{mv^2}{r}$$

$$= \frac{1500\text{Kg} \times (20)^2\text{m/s}}{20\text{m}} = \frac{1500 \times 20 \times \cancel{20}}{\cancel{20}}$$

$$F_c = 1500 \times 20 = 30000\ \text{N}$$

7.5 CENTRIFUGAL FORCE:

When a body is moving along circular path it has a tendency to move along the tangent but a force is acting on it directed towards the centre of circle which keeps the body to move in circle called centripetal force. According to Newton's third law of motion, there must be a reaction force directed away from the centre of circle and is equal in magnitude but opposite to centripetal force called centrifugal force.

When a stone at the end of a string whirl, to keep the stone in circular motion a force act pulls it inward along the string. According to third law of motion stone reacts and exert an equal force outward on the hand. The outward force on the hand is called centrifugal force.

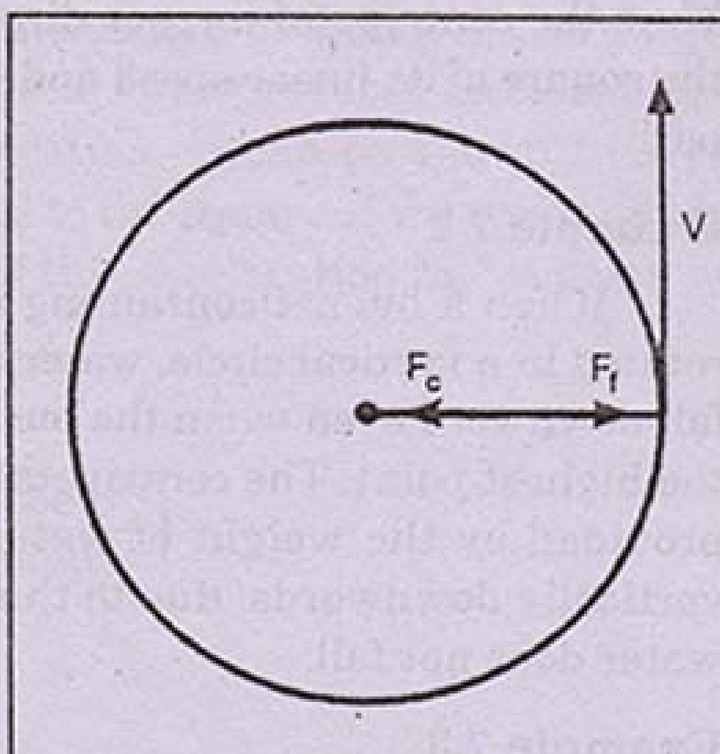


Fig. 7.5

7.6 BANKING OF THE ROADS:

When a moving car turns round a corner the man (occupents) inside the car tends to fall outward direction. To prevent them from falling sufficient centripetal force is required. On level curve (unbanked) when car is going, three forces are acting on it, it's weight vertically downward, the upward push of the road and force of friction between road and tires which provide necessary centripetal force to cause the car to maintain the acceleration for normal speed. If the speed of car exceeds a certain limit, there will be insufficient force to cause the car to maintain acceleration and the car is unable to go around the curve and will skid in a curve of greater radius.

To provide sufficient force the curves on highways are banked, computed to be safe at some particular speed. Thus car going properly on banked curve at some particular speed. When the force exerted by the road perpendicular to it's surface has a horizontal component equal to the required centripetal force and vertical component equal and opposite to the weight of car (vehicle). Than car easily going round the banked road.

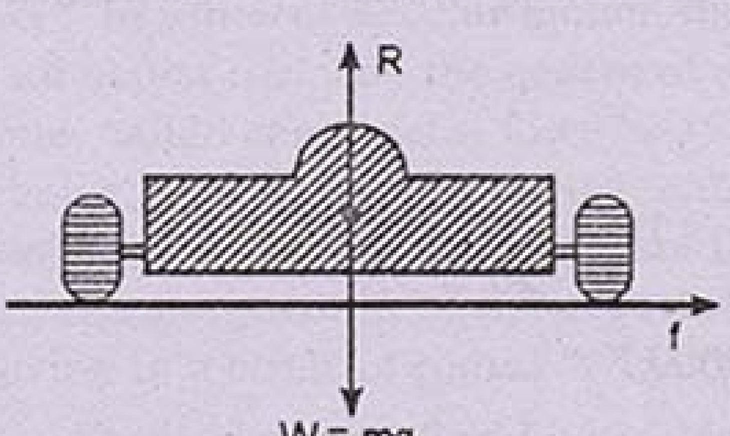


Fig. 7.6

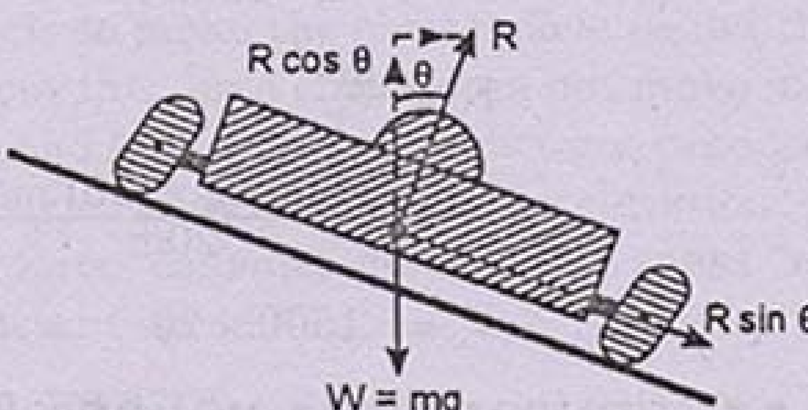


Fig. 7.7

7.7 CENTRIFUGE:

It is an appliance used to seperate heavier particles from the lighter particle in a liquid. The liquid is rotated in a cylindrical vessel at high speed

with the help of an electric-motor. The heavier particle moves away from the axis of rotation and lighter particals moves near to the axis of rotation.

(i) In washing machine dryer the wet clothes are rotated at high speed. The water particals due to centrifugal force throws outword through the holes in the wall of outer vessal as the drum containing wet clothes rotates.

(ii) A cream seperator is a type of centrifuge in which milk is rotated at high speed, the lighter cream particals collect near the axis while the skimmed milk moves away from axis.

(iii) Sugar crystals are seperated from molasses with the help of centrifuge when sugar solution is rotate in a cylindrical vessal the sugar crystals moves away from the axis of rotation and collected on the net around the vessal drum.

7.8 LAW OF UNIVERSAL GRAVITATION:

The Law of Universal Gravitation was discovered by Sir Isaac Newton in the year 1686. According to Newton, this law is true not only for the heavenly bodies but is also true of any two bodies in this universe. According to this law "Every body in this universe attract every other body with a force which is directly proportional to the product of their masses and inversely proportional to the square of distance between their centres".

Consider two bodies "A" and "B" of masses 'm_1' and 'm_2' respectively that are placed with their centres at distance "r" from each other as shown in the Fig. 7.8.

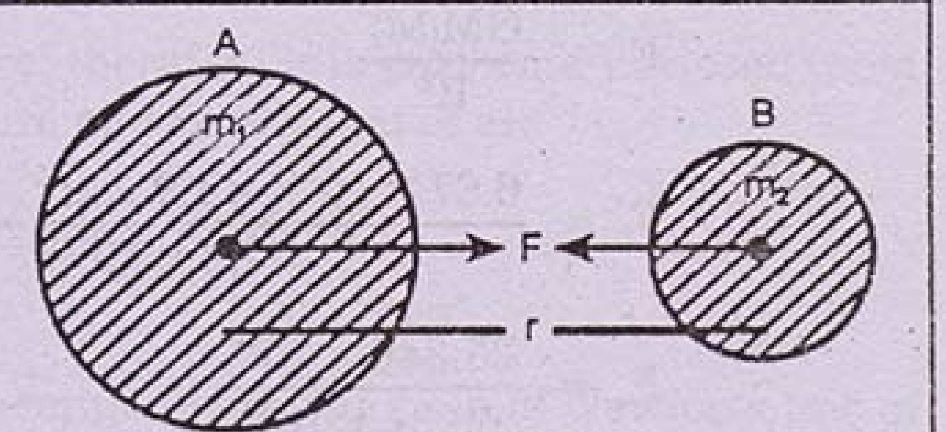


Fig. 7.8

According to Newton's law of gravitation the force of attraction exerted by body 'A' on 'B' body is same in magnitude but opposite in direction to the force which exerted by body 'B' on 'A' body.

$$F \propto m_1 m_2 \text{ .... (i)}$$

$$F \propto \frac{1}{r^2} \text{ .... (ii)}$$

Combining (i) and (ii) then $F \propto \frac{m_1 m_2}{r^2}$ .... (iii)

The proportion above may be converted to an equation on multiplication by a constant "G" which is called the gravitational constant

$$F = \frac{Gm_1 m_2}{r^2} \text{ .... (7.3)}$$

The numerical value of constant "G" depends on the units in which force, mass and distance are expressed. In S.I. Units the value of

$$G = 6.67 \times 10^{-11}\ \text{N-m}^2/\text{Kg}^2$$

$$\text{or}\quad G = 6.67 \times 10^{-8}\ \text{dyne-m}^2/\text{Kg}^2$$

Example 7.5

The mass of the planet Jupiter is 1.9×10^{27} Kg and that of sun is 2×10^{30} Kg. If the average distance between them is 7.8×10^{11} m, find the gravitational force of Sun on Jupiter.

Solution:

M_j = Mass of Jupitar = 1.9×10^{27} Kg

M_s = Mass of Sun = 2×10^{30} Kg

R = Distance between the center of Sun and Jupiter

$$R = 7.8 \times 10^{11}\ \text{m}$$

$$F = \frac{GM_jM_s}{R^2}$$

$$F = \frac{6.67 \times 10^{-11} \times 1.9 \times 10^{27} \times 2 \times 10^{30}}{(7.8 \times 10^{11})^2}$$

$$F = \frac{25.346 \times 10^{46}}{60.84 \times 10^{22}}$$

$$F = 0.4166 \times 10^{24}\ \text{N}$$

$$F = 4.166 \times 10^{23}\ \text{N}$$

7.9 GRAVITATIONAL CONSTANT "G"

The value of "G" was determined experimentally by Henry Cavendish in 1798 with an instrument called Cavendish balance.

The Cavendish balance consists of two small spheres of mass 'm' mounted at opposite ends of a light horizontal rod of length 'd' which is suspended at it's centre by a fine vertical fibre. When two large spheres of mass 'M' are brought near to small spheres at distance "r", as shown in Fig. The force of gravitational attraction between the spheres produces torque, which twists the fibre. The twist in fibre is proportional to the magnitude of torque.

If "$\frac{GmM}{r^2}$" is the gravitational force between the spheres and "$\frac{GmM}{r^2} \times d$" is the torque which produce twist "θ" in the fibre. Than

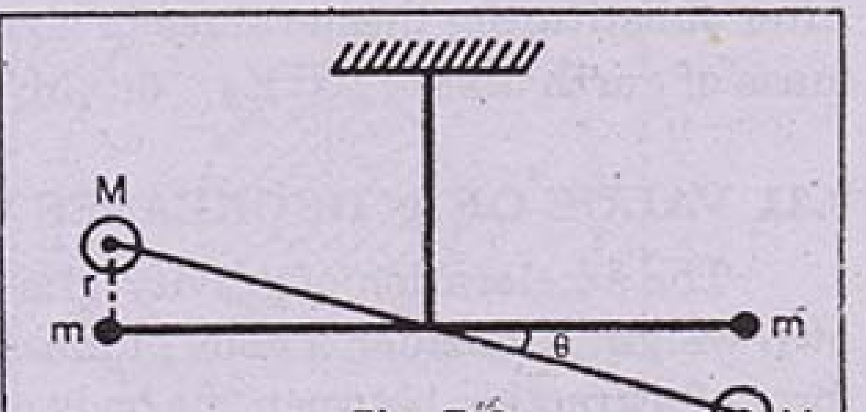


Fig. 7.9

$$\frac{GmM}{r^2} \times d \propto \theta$$

or $$\frac{GmM}{r^2} \times d = C\theta$$

Where 'C' is a proportionality constant. It's value depends upon the nature of the material of fibre. For different material it's value is different it can can be determined experimentally.

By substituting the values of all known quantities in the above equation, the value of "G" can be calculated.

$$\therefore \quad G = \frac{r^2}{Mmd} C\theta \text{ ---------- (7.4)}$$

This experiment needs very care, therefore it is performd in closed room so as to prevent change in temperature and circulation of air.

The accepted value of "G" = 6.67×10^{-11} N-m^2/Kg^2

7.10 MASS OF EARTH

The mass of earth may be determined with the help of Newton's law of universal gravitation. We know that the earth attract the bodies towards it's centre. Consider a body of mass 'm' placed on the surface of earth. If 'M_e' be the mass of the earth and 'R_e' it's radius than the force with which the earth attract the body towards it's centre is given by equation.

$$F = \frac{GmM_e}{r_e^2}$$

But the force of attraction is equal to the weight of the body

$$\therefore \quad W = \frac{GmM_e}{r_e^2}$$

or $$mg = \frac{GmM_e}{r_e^2}$$ or $$g = \frac{GmM_e}{r^2}$$

or $$M_e = \frac{gR_e^2}{G} \text{ ------------ (7.5)}$$

Since $g = 9.8 m/sec^2$, $\quad G = 6.67\times10^{-11}\ N\text{-}m^2/Kg^2$

and $\ R_e = 6.38 \times 10^6 m$

After substituting these values in above equation, we find the mass of earth $5.98 \times 10^{24} Kg$ or $M_e = 6 \times 10^{24} Kg$

7.11 VALUE OF 'g' DECREASES WITH ALTITUDE:

The acceleration of gravity 'g' is the acceleration imparted to a body by it's own weight. Consider a body of mass 'm' on earth's surface. The gravitational force of attraction between the body and earth is equal to the weight of the body

$$F = W = \frac{GmM_e}{R^2}$$

or $$mg = \frac{GmM_e}{R^2}$$

or $$g = \frac{GM_e}{R^2} \quad \text{--------------} \quad (7.6)$$

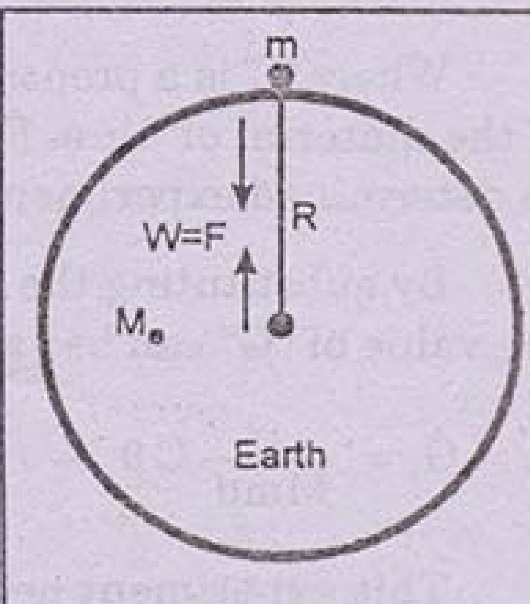


Fig. 7.10

Since "G" and "M_e" are constant, the acceleration of gravity "g" decreases with increasing distance "R" from the centre of earth

$$g \propto \frac{1}{R^2}$$

The "g" at any point above the surface of earth will vary inversely as the square of the distance from that point to it's centre.

If body is moved up to "h" distance from earth surface the acceleration due to gravity at that point is

$$g_h = \frac{GM_e}{(R+h)^2} \quad \text{------------------} \quad (7.7)$$

Divide equation (7.7) by (7.6), we get

$$\frac{g_h}{g} = \frac{GM_e}{(R+h)^2} \div \frac{GM_e}{R^2}$$

or $$\frac{g_h}{g} = \frac{GM_e}{(R+h)^2} \times \frac{R^2}{GM_e} = \frac{R^2}{(R+h)^2}$$

$$\frac{g_h}{g} = \frac{1}{\frac{(R+h)^2}{R^2}} = \frac{1}{1+\frac{2h}{R}+\frac{h^2}{R^2}}$$

As $\frac{h^2}{R^2}$ is negligibly small

$$\frac{g_h}{g} = \frac{1}{1+\frac{2h}{R}} = \left(1+\frac{2h}{R}\right)^{-1}$$

or $$g_h = g\left(1-\frac{2h}{R}\right) \cdots (7.8)$$ The terms containing higher powers of $\frac{h}{R}$ are neglected.

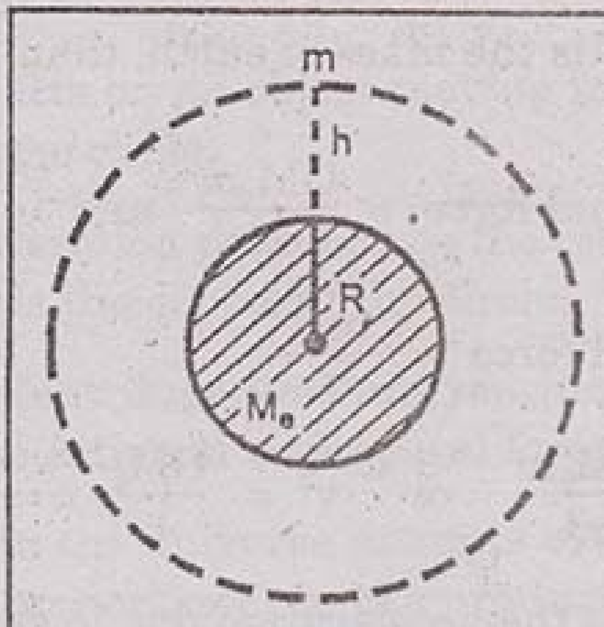


Fig. 7.11

This equation show that greater the value of 'h' the smaller is the value of 'g_h' or value of 'g' decreases with altitude.

7.12 ARTIFICIAL SATELLITES

In the universe, planets revolve around the sun in different orbits with different speed and time period. Similarly the moon revolves around the earth. The force of gravity between planets and sun or between moon and earth provides the necessary centripetal force for them to be revolving in their orbits. Hence moon is the satellite of the earth. Many artificial satellites for different purposes were launched around the earth with the help of multistage rockets since 1957.

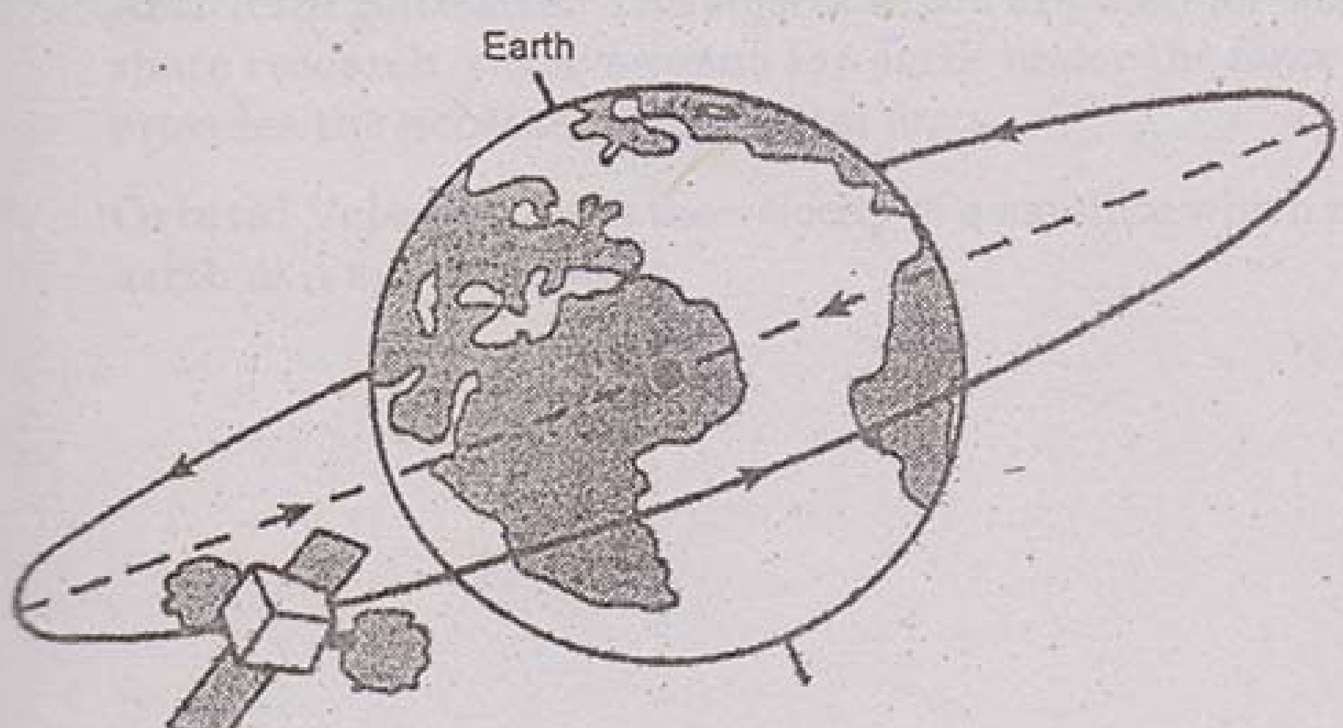


Fig. 7.12

7.13 ORBITAL VELOCITY:

Consider a satellite of mass 'm' moving in an orbit of radius 'r' with velocity v around the earth. The gravitational force of attraction between the satellite and earth provides necessary centripetal force

If 'M_e' is the mass of earth, than

$$\text{Gravitational force} = \frac{GM_em}{r^2} \quad \text{and}$$

$$\text{Centripetal force} = \frac{mv^2}{r}$$

$$\frac{mv^2}{r} = \frac{GM_em}{r^2} \quad \text{or} \quad v^2 = \frac{GM_e}{r} \quad \text{............ (i)}$$

If 'g' is the acceleration due to gravity at the position of satellite

$$\text{then} \quad mg = \frac{GM_em}{r^2} \quad \text{or} \quad gr = \frac{GM_e}{r} \quad \text{............ (ii)}$$

Compare equation (i) and (ii) therefore; we have

$v^2 = gr$ or $v = \sqrt{gr}$

If satellite, revolving very near to the surface of the earth

then $r = R_e$ = radius of the earth

$\therefore \ v = \sqrt{gR_e}$

$\therefore \ v = 7920$ km/s

SUMMARY

1. **Centripetal Force:** A force, which acts on an object moving in a circle and is directed towards the centre of the circle.

2. **Centripetal Acceleration:** The acceleration produced by the centripetal force, which is always directed towards the centre of the circle.

3. **Application of Centripetal Force:** Centrifuge, washing machine, cream separator are the devices based on the action of centripetal force.

4. **Law of Gravitation:** Every object in the universe attracts every other object with a force which is directly proportional to the product of their masses and inversely proportional to the square of the distance between their centres.

$$F = G\,\frac{m_1 m_2}{d_2}$$

5. **Variation of 'g' with altitude:** The value of g decreases as altitude increases given by

$$g_h = \frac{gR^2}{(R+h)^2}$$

6. **Artificial Satellite:** The objects which are used for communication and space research, revolve round the earth under the force of gravity, which provides the necessary centripetal force.

7. **Orbital Velocity:** It is the velocity of a satellite which moves around the earth at a specific height.

$$V_{ort} = \sqrt{\frac{GM}{R}}$$

QUESTIONS:

7.1 Answer the following questions:

1. Is it possible for a body to be accelerated if it's speed is constant?
2. Explain what is meant by centripetal acceleration.
3. What is centripetal force. Give examples of a body moving in circular path.
4. Why is the acceleration of a body moving uniformly in a circle, directed towards the centre?
5. When a moving car turns round a corner to the left, in what direction do the acceptants tends to fall?
6. What is a centrifuge. Give some practical applications of a centrifuge.
7. State and explain Newton's law of gravitation.
8. How one can be determine the mass of earth.
9. How would the value of 'g' and 'G' be effected, if the mass of earth becomes four times?
10. Explain how the value of 'g' decreases with a change in the altitude.

7.2 Fill in the blanks.

1. 'G' is called __________.
   (a) Gravitational acceleration
   (b) Gravitational force
   (c) Gravitational constant
2. In case of satellites the necessary acceleration is provided by __________.
   (a) Frictional force (b) Gravitational force
   (c) Coulombs force
3. The centripetal force is always directed to __________.
   (a) Towards the centre of circle
   (b) Along the direction of motion
   (c) Away from centre of circle
4. If the speed of body moving in circle is doubled it's centripetal acceleration becomes __________.
   (a) Twice (b) Four times
   (c) Eight time

7.3 Pick out true and false from the following:

1. Law of universal gravitation is applicable not only to heavenly bodies but also for small objects.
2. A body moving with uniform speed on a circle has a uniform acceleration.
3. In centrifuge machine heavier particles moves towards the axis of rotation.
4. The value of "G" have same value at Karachi and on Mount Everest.

PROBLEMS

7.1 At what speed must an object travel in a circle of radius 2m to experienc a centripetal acceleration of $9.8m/s^2$.

(4.427 m/s

7.2 A car goes round a curve at 20m/s. The radius of curvature is 50m. Calculat the centripetal acceleration of the car.

($8m/s^2$

7.3 A proton of mass 1.67×10^{-27}kg is moving in a circle of radius 100cm. Ar electro-magnet applies a force of 1×10^{-12}N directed towards the centre o the circle. What is the velocity of the proton.

(2.44×10^7m/s

7.4 A string 2m long is used to whirl a 200gm stone in horizontal circle at speed of 2m/s. Find tension in string.

(0.4N

7.5 Compute the gravitational force of attraction between two boys of masse 50kg and 40kg respectively apart from each other.

(13.34×10^{-8}N

7.6 Compute the force of gravitation between the small and large spheres o Cavendish balance if m=1gm and M=300gm & r=5cm.

(8×10^{-7} dyne cm^2/gm^2

7.7 Two balls of 40kg and 20kg masses attract each other with a force o 3.33×10^{-7}N. Find the distance between the masses if the value o $G=6.67\times10^{-11}N\text{-}m^2/kg^2$.

(0.4m

7.8 The distance from the earth to sun is 1.49×10^{11}m. The mass of earth i 6×10^{24}kg and mass of sun is 2×10^{30}kg compute the gravitational forc between them.

(3.6×10^{22}N

CHAPTER – 8

WORK, POWER AND ENERGY

LEARNING OBJECTIVES:

- Introduction.
- Work.
- Power.
- Energy.
- Kinetic energy (K.E).
- Gravitational potential energy.
- Elastic potential energy.
- Interconversion of kinetic energy and potential energy.
- Conservation of energy.

8.1 INTRODUCTION

The word work is a familiar one, used in an ordinary sense. In physics, however, it is used in a restricted and carefully defined manner. A person does work while lifting an object from floor to roof; a horse does work when pulling a cart. In physics work is done only if there is displacement of a body on which force acts.

Energy is the most important physical concept which is studied in all sciences. Work is closely related with energy which provides a link between force and energy. Work, energy and power have special meaning in physics. The concepts and applications of work, power and energy will be discussed in this chapter.

8.2 WORK

When force acts on a body and the body moves through some distance along the direction of force, then work is said to be done. If a constant force 'F' acts on a body placed on a horizontal surface and displacement 'd' takes place in it along the direction of force, then the work done by this force is defined as the product of the magnitude of force and that of the displacement:

$$W = (F)(d) = Fd \quad \text{———} \quad (8.1)$$

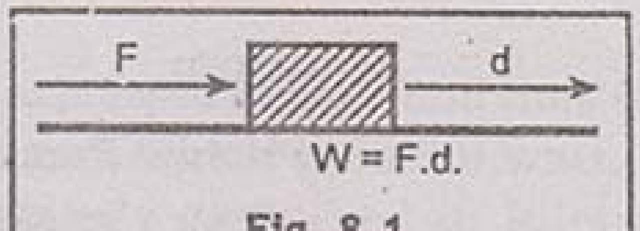


Fig. 8.1

If a constant force '$\vec{F}$' is acting on the body at an angle 'θ' with the direction of motion of the body then work is defined as the product of the magnitude of the displacement and of the component of the force in the direction of displacement. (Fig. 8.2)

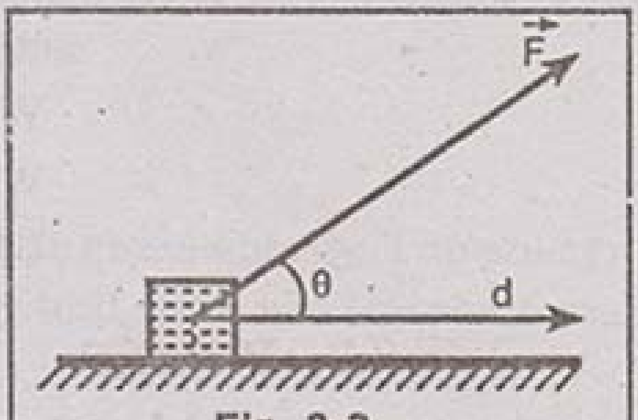


Fig. 8.2

$$W = (F\cos\theta)\, d.$$

or $\quad W = F(d\cos\theta)$

Where ($\vec{F}\cos\theta$) is the magnitude of the component of '$\vec{F}$' in the direction of 'd' and ($d\cos\theta$) is the component of displacement in the direction of '$\vec{F}$'.

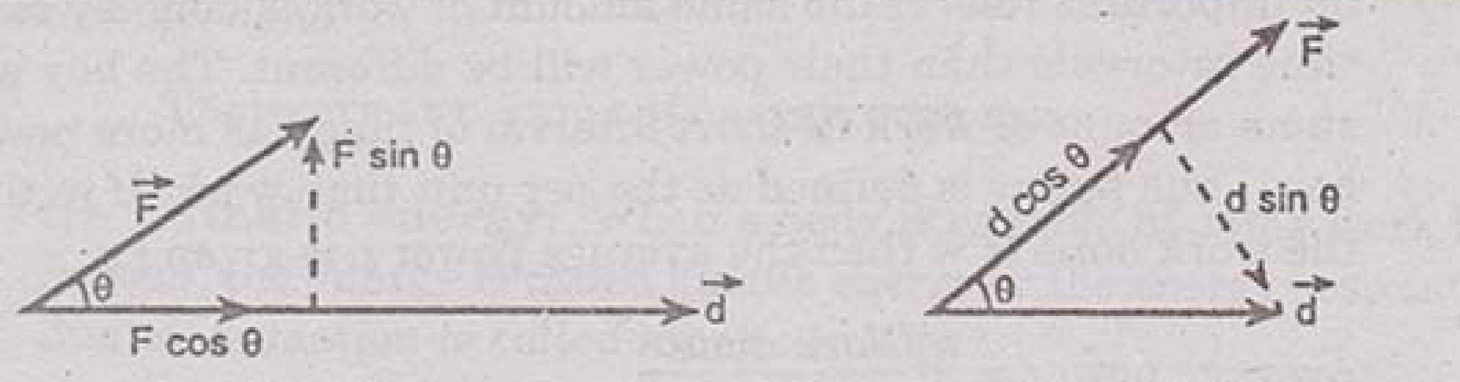


Fig. 8.2 (a) Fig. 8.2 (b)

If the force is in the same direction as the displacement, the work is positive. Here θ =0°, the work will be positive and maximum. If the force is opposite to the displacement the work is negative, and if force is at right angle to the displacement, it has no component in the direction of displacement and the work will be zero.

In the International system the unit of force is Newton and the unit of distance is metre. Hence the unit of work is newton-metre which is called joule. In S.I Units the unit of work is joule. 1 joule = 1N × 1m

1 joule = 1Nm

Joule is defined as the work done by a force of one newton in moving a body through a distance of one meter in the direction of force.

Example 8.1

When a force of 100N making an angle of 45° with the horizontal displaces a body through 20m. Find the work done.

Solution:

F = Force applied = 100N.

θ = angle with horizontal = 45°.

d = distance covered = 20m.

Formula

$$W = Fd\cos\theta$$

$$W = 100N \times 20m \times \cos 45° \text{ N–m}$$

$$W = 2000 \times .707 \text{ N–m}$$

$$W = 1414 \text{ joules.}$$

8.3 POWER

In the definition of work no time factor is involved but in power, time pla an important role. If the same amount of work is done by two boys in unequ time intervals then their power will be different. The boy who completes t same amount of work in short interval of time has more power than the oth boy. Thus power is defined as the per unit time work. If in the time interval the work done is w then the average power p is given as

$$\text{Average power} = \frac{\text{Work done}}{\text{time interval}}$$

$$P = \frac{W}{t} \quad \text{———————— 8.2(a)}$$

Power and Velocity:

Suppose a constant force 'F' acts on a body and displaces it through distar "d" in the direction of force in time "t", the work done is

$$W = (F)(d) = Fd$$

and average power developed is

$$P = \frac{W}{t}$$

$$P = \frac{Fd}{t} = F\frac{d}{t} \quad \text{but}$$

$$\frac{d}{t} = \text{average velocity}$$

Hence $P = Fv$ ———————— 8.2(b)

Hence power is the product of the force and velocity.

In the International system work is expressed in joule and time in second Therefore the unit of power is joule/second, which is called watt. This is small unit and so power is more commonly expressed in kilowatts(kw 1Kw = 1000 watts.

In the British engineering system the unit of power is horse power (hp where 1hp = 550 ft lb/s and 1hp = 746 watts.

8.4 ENERGY

Whenever work is done on an object, it gains energy. The energy of a boc is it's capacity of doing work. Thus energy and work are the same and both a measured in the same units.

Whenever we do work on a body, we store in it an amount of energy equ to work done.

There are two forms of energy, such as:

(i) **Kinetic energy:** When a body is capable of doing work by virtue of it motion, the energy is called kinetic energy or the energy associated with a ma due to its motion is called kinetic energy.

(ii) **Potential energy:** When a body is capable of doing work by virtue of it position, or special configuration the energy is called potential energy.

There are different kinds of potential energy such as:

(1) **Gravitational P.E:** The energy associated with a mass when lifted to height in the earths gravitational field is called gravitational potential energ

(2) **Elastic Potential energy:** The energy stored in a stretched or compresse elastic material such as a spring is called elastic potential energy.

8.5 KINETIC ENERGY (K.E)

To obtain an expression for K.E we have to determine the work done b the body in motion. This work is equal to the Kinetic energy of the body. Consid a body of mass 'm' placed on a horizontal surface initially at rest. When a for 'F' is applied, it covers a distance 'S' and it's final velocity becomes "v". Workdor is covering the displacement S.

$$W = F.S \text{ ———— (i)}$$

But by second law of motion, when a force acts on a body, it produces an acceleration in the direction of force

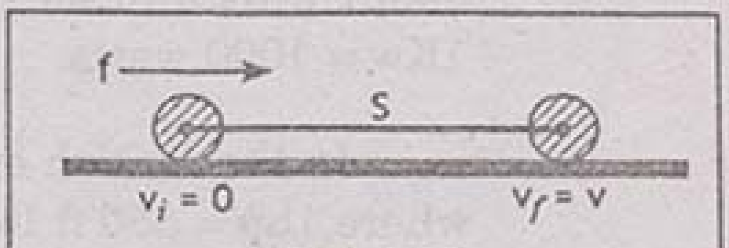


Fig. 8.3

$$F = ma \text{ ———— (ii)}$$

And by using the equation of motion

i.e. $v_f^2 - v_i^2 = 2aS$

When $v_i = 0$, $v_f = v$, $s = ?$

$$\therefore \; v^2 - 0 = 2as \text{ or } s = \frac{v^2}{2a} \text{ ———— (iii)}$$

Putting the value of 'F' and 's' in equation(i)

$$\text{Work} = K.E = F.S = ma \times \frac{v^2}{2a}$$

$$\therefore K.E = \frac{mv^2}{2} = \frac{1}{2} mv^2$$

$$K.E = \frac{1}{2} mv^2 \text{ ———— (8.3)}$$

The equation shows that the Kinetic energy of a body is directly proportional to it's mass and also to the square of it's velocity. If two locomotive move with same velocity then the K.E of the locomotive which has large mass is greater than the other. If, however, two cars of same mass move with different velocities, then the one having greater velocity will possess greater kinetic energy.

8.6 GRAVITATIONAL POTENTIAL ENERGY

The potential energy possessed by a body in the gravitational fields called the gravitational P.E.

If a body is lifted up to a position higher than its initial position in the gravitational field of earth, work is done on it. This work is stored in the body as potential energy. Such potential energy is called gravitational potential energy (P.E).

Consider a body of mass 'm' on the ground. If it is lifted up with constant speed through a vertical height "h", the force required to raise the body is just equal and opposite to its weight W=mg. Thus work done on it against gravitational field is store in it as gravitional work = mgh

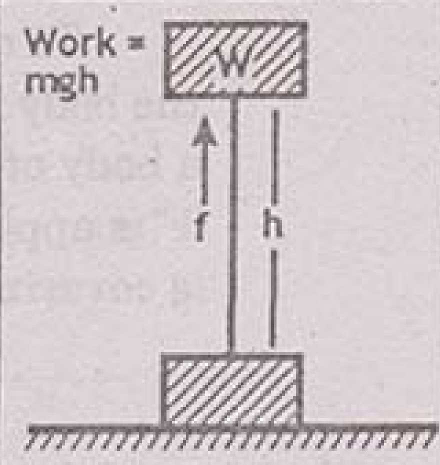


Fig. 8.4

Potential energy (P.E)

$$\therefore P.E = W.h = mgh \text{ ———— (8.4)}$$

Example 8.2

A car weighing 19600N is moving with a speed of 30 m/sec on a level road. If it is brought to rest in a distance of 100 m, find average frictional force acting on it.

Solution:

The car possesses K.E. which is used up in doing work against the force of friction.

$\therefore$ K.E = Work done against force of friction

$$\frac{1}{2}mv^2 = f.d.$$

$$\frac{1}{2} \times \frac{19600N}{9.8} \times 30 \text{ m/s} \times 30 \text{ m/s} = f \times 100m$$

$$f = \frac{1000 \times 900}{100} = 9000 \text{ N}$$

Here 'f' is the frictional force.

8.7 ELASTIC POTENTIAL ENERGY

If a spring is wound, work is done in winding it against the elastic force of spring. This work is stored in the spring as potential energy. This type of potential energy is due to constrained position of the spring and is called elastic potential energy.

To find the energy stored in a compressed spring, we calculate the work required to compress it. In figure 8.5(a) a force 'F' pushes a spring to compress it from its equilibrium position 'O' to some other position x, this force is directly proportional to the amount of the compression.

i.e. $F \propto x$ or $F = Kx$ ———— (i)

Where 'K' is force constant (spring constant) its value depends on how stiff the spring is. Since the compression force is zero at 'O' and Kx at x, the average force needed to compress the spring from position O to x is

$$F_{av} = \frac{O + Kx}{2} = \frac{Kx}{2} = \frac{1}{2}Kx$$

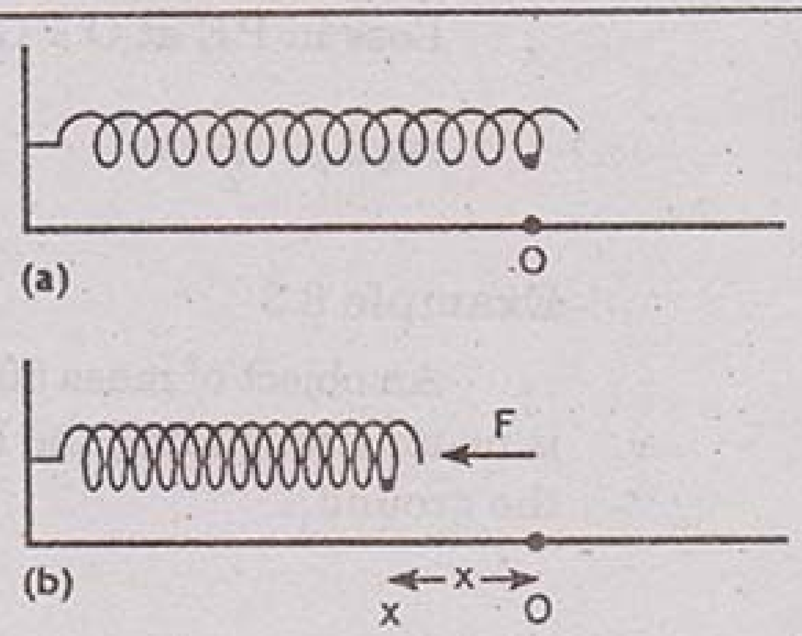


Fig. 8.5

Therefore work done in compressing the spring is $W = F_{av}x$

or $\quad W = \frac{1}{2} Kx \times x \;=\; \frac{1}{2} Kx^2$

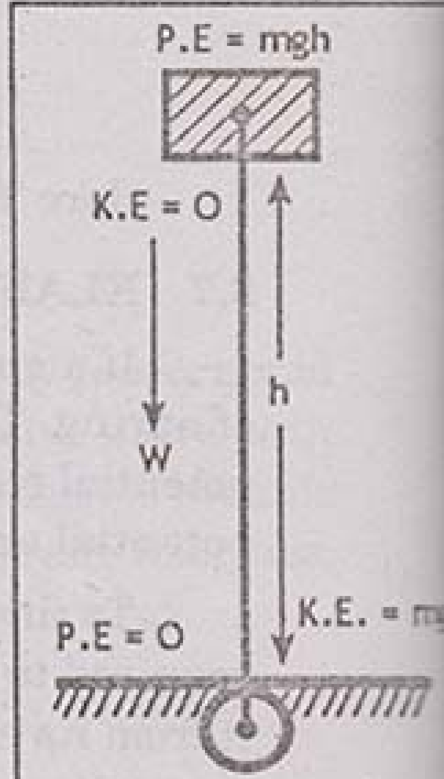
This work causes the elastic potential energy, therefore

$$\text{Elastic Potential Energy} = \frac{1}{2} Kx^2 \text{ ——————————— } (8.6)$$

8.8 INTERCONVERSION OF KINETIC ENERGY AND POTENTIA ENERGY

Scientists believe that the amount of certain quantities in the univers never changes. These quantities are called conserved quantities. Important ar useful aspect of energy is the fact that energy is a conserved quantity and th total amount of energy always remains constant.

Consider an object of mass 'm' lying at a height 'h' from earth's surface at point 'P'. It possesses potential energy equal to mgh and Kinetic energy equal to zero. Let the body be allowed to fall down under the action of gravity. During its downward motion, its height above the ground decreases continuously so its potential energy also decreases. As the body has been accelerating under the action of gravity, hence the work done by it is equal to its Kinetic energy. Thus when the body falls freely under gravity and when there is no opposing force there is a continuous decrease in its P.E and this decrease is equal to increase in K.E. When it reaches the surface of the earth its P.E. is zero and whole of the energy is Kinetic.



(Fig. 8.6)

Loss in P.E at O = Gain in K.E. at O.

$$mgh = \frac{1}{2} mv^2$$

Example 8.3

An object of mass 500 gm falls from a height of 20m on the ground, if the is no air friction find the Kinetic Energy possesed by object just before touchi the ground.

Solution:

Potential energy at a height of 20m = Kinetic energy just before touching the ground.

Loss in P.E. = Gain in K.E.

$$mgh = \frac{1}{2}mv^2$$

$$0.5Kg \times 9.8 \times 20m = K.E.$$

$$9.8 \times 10 = K.E.$$

$$98 = K.E.$$

$$\therefore \ K.E. = 98 \text{ joules}$$

8.9 CONSERVATION OF ENERGY:

Energy is a conserved quantity. The amount of energy in the universe is fixed. Energy can be transferred between system and the environment and can also be converted from one form to another. There are many forms of energy. All forms of energy are exchangeable with mechanical energy. In section 8.8 we have seen that in the absence of any opposing force, the loss of potential energy of a body is equal to the gain in its kinetic energy; this means the total energy of the body remains constant. We have many forms of energy such as light energy, mechanical energy, electrical energy, heat energy etc. Whenever energy of some kind disappears, an equivalent amount of other form of energy appears. In this process the total amount of energy in the universe remains constant. This fact explains a general law known as law of conservation of energy. It states that "Energy can neither be created nor destroyed, it can be changed from one form to an other form but the total amount of energy remains constant".

SUMMARY

1. **Work:** When a body covers some distance in the direction of the force then work is said to be done. The product of force and the displacement covered in the direction of the force is called work.

2. **Unit of work:** Work will be one joule if a force of one newton moves a body through a distance of one metre in the direction of the force.

3. **Power:** Rate of doing work with respect to time is called power.

4. **Unit of Power:** Power will be one watt, if one joule work is done in one second.

5. **Energy:** Ability of a body to do work is called energy. Thus, its unit is also a joule.

6. **Kinetic Energy:** Ability of a body to do work due to its motion is called kinetic energy.

7. **Gravitational Potential Energy:** Energy of a body due to its height from the surface of the earth is called gravitational potential energy.

8. **Elastic Potential Energy:** Energy of a stretched or compressed spring is called elastic potential energy.

9. **Law of Conservation of Energy:** One form of energy can be converted into another form of energy but total amount of energy remains constant.

QUESTIONS

8.1 Answer the following questions.

(i) Define work, positive and negative work and the unit of work.

(ii) What is energy, name the different forms of energy?

(iii) What happens to the potential energy of a body when dropped from certain height.

(iv) Does a hydrogen filled balloon possesses any potential energy? explain.

(v) Is energy a vector quantity?

8.2 Fill in the blanks.

(i) The work will be maximum if the angle between force and displacement is ___________.

(ii) If the angle between force and displacement is 180° then work done will be __________.

(iii) Power is the product of ________ and velocity.

(iv) The ability of a body to do work is called ________.

(v) 1hp = _________ watts.

(vi) Energy is a _________ quantity.

8.3 Identify the correct answer.

(i) 1 Joule is equal to.

(a) 1N-S. (b) 1lb-m. (c) 1N-m.

(ii) Power is defined as.

(a) Rate of change of position.

(b) Rate of change of force. (c) Time rate of doing work.

(iii) The energy possessed by a body due to its position is called:

(a) Kinetic energy. (b) Heat energy.

(c) Potential energy.

(iv) The work will be positive, if the angle between force and displacement is

(a) 90°. (b) 180°. (c) 0°.

PROBLEMS

3.1 How much work is done to displace horizontally a body 40m by a force of 200N, whose angle with the horizontal is 30°?

(6928J)

3.2 What is the power of an engine that pulls a 1000 Kg automobile at a steady speed of 10 m/s along a level road?

(98000 watt)

3.3 With what constant velocity can a 1960 watt motor raise a mass of 100 Kg.

(2m/s)

3.4 A boy of mass 50Kg on motor bike is moving with 20 m/s. What is his K.E?

(10000J)

3.5 What is the P.E stored by a car of weight 2000N when lifted 50m straight up.

(100000J)

3.6 A box is pushed 5m across a level surface by a horizontal force of 200N. How much work is done on the box?

(1000J)

3.7 What is the K.E. of 2000 Kg car travelling at 90 Km/h.

(625000J)

3.8 An object weighing 10N falls through a distance of 10m. What is its velocity just before it strikes the ground. Assume there is no friction?

(14 m/s)

CHAPTER – 9

SIMPLE MACHINES

LEARNING OBJECTIVES:

- Simple machine.
- Mechanical advantage.
- Kinds of simple machines.

9.1 SIMPLE MACHINE

A machine is a device which is used to do work in a more convenient way or used to speed up work.

A machine can also be defined as any device by means of which a force applied at one point can be used to overcome a resistance at some other point.

Machines are used for various purposes. They are as follows:

(1) TRANSFER OF FORCE:

With the help of machine we can transfer force from one place to another. e.g.,

(a) While driving a nail into a wall, the force of our hand is transferred to the nail with the help of hammer.

(b) While driving a cycle we transfer the force of our feet to its wheels.

(2) ENLARGEMENT OF THE EFFECT OF FORCE:

With the help of a machine we can lift much heavier load or overcome much greater resistance by applying comparatively smaller force. e.g., we lift heavy vehicles with the help of a screw jack. In this case the screw jack enlarges the magnitude of force of our hand.

(3) CHANGE IN THE SPEED OF WORK:

Some machines are used to increase the speed of work e.g., by using a bicycle we can travel larger distance then by foot in the same interval of time, while in some cases machines are used to reduce the speed of work e.g., use of gears in a car.

(4) CHANGE IN THE DIRECTION OF FORCE:

Some machines are used to change the direction of the force e.g., while raising or lifting the bucket of water from a well we use a pulley and in this case we apply the force in the downward direction while the bucket moves in the upward direction.

There are certain terms which are used in the study of machines. These terms are given below:

(i) Effort:

The force applied to a machine for doing work is called effort. It is denoted by "P".

(ii) Load:

The weight lifted or resistance overcome by a machine is called load and is denoted by "W".

(iii) Input:

The work done on a machine by the effort is called input. If an effort P acts through a distance d, then the work done by the machine or input is given by.

Input = Effort × distance through which the effort acts

$$\text{Input} = P \times d \quad \text{———} \quad (9.1)$$

(iv) Output:

The work done by the machine on the weight is called output. If a machine moves a load "W" through a distance "h" then the work done or the output of the machine is given by

Output = Load × distance through which the load moves

$$\text{Output} = W \times h \quad \text{———} \quad (9.2)$$

9.2 MECHANICAL ADVANTAGE:

The ratio between the load lifted and the effort applied is called the mechanical advantage of a machine.

If W is the weight lifted by the machine and P is the effort applied to the machine then

$$\text{Mechanical Advantage} = \frac{\text{Weight lifted by the machine}}{\text{Effort applied}}$$

$$\text{M.A.} = \frac{W}{P} \quad \text{———} \quad (9.3)$$

Since mechanical advantage is a ratio between two forces, so it has no unit. It is expressed in numbers.

EFFICIENCY:

The ratio between the useful work done by the machine (output) to the work done on the machine (input) is called efficiency.

$$\text{Efficiency} = \frac{\text{Useful work done by the machine}}{\text{Work done on the machine}}$$

$$= \left(\frac{\text{Output}}{\text{Input}}\right)$$

$$= \left(\frac{Wh}{Pd}\right) \quad \text{———} \quad (9.4)$$

Efficiency is usually expressed in percentage.

$$\text{Eff} = \frac{Wh}{Pd} \times \frac{100}{100}$$

$$= \frac{100\,Wh}{Pd}\,\%$$

The efficiency of a real machine is always less than 1. A perfect machine has 100% efficiency or 1.

For an ideal machine

Efficiency = 1

Or $\frac{\text{Output}}{\text{Input}} = 1$

Or Output = Input

Example 9.1

Calculate the mechanical advantage of a crow-bar which can lift a weight 100 N on the application of 50 N force.

Solution:

Effort applied = P = 50 N
Weight lifted = W = 100 N
Mechanical advantage = ?

$$\text{Mechanical Advantage} = \frac{\text{Weight}}{\text{Effort}}$$

$$= \frac{100}{50}$$

$$= 2$$

Example 9.2

600 joules of energy is supplied to a machine to lift a load of 300 N through a vertical height of 1 m. Find the efficiency of the machine.

Solution:

Input = 600 J

Weight lifted = 300 N

Distance = 1 m

Output = Weight × distance covered by weight

= 300 × 1

= 300 J

$$\text{Efficiency of a machine} = \left(\frac{\text{Output}}{\text{Input}}\right) \times 100$$

$$= \left(\frac{300}{600}\right) \times \frac{100}{100}$$

$$= 50\%$$

9.3 KINDS OF SIMPLE MACHINES

In our daily life we use different kinds of machines. All the machines, however complicated, may consist of one or more than one of the following six simple machines.

(1) Lever.

(2) Inclined Plane.

(3) Pulley.

(4) Screw Jack.

(5) Wedge.

(6) Wheel and Axle.

LEVER

A lever is one of the simplest machines known. It is used to lift heavy loads by the application of small force (effort).

"A lever is a rigid bar which rotates about a fixed point called fulcrum. By applying force at one end of a bar, weight can be lifted at the other end. The perpendicular distance between force and fulcrum is called force arm, or effort arm and the perpendicular distance between fulcrum and weight is called weight arm or load arm.

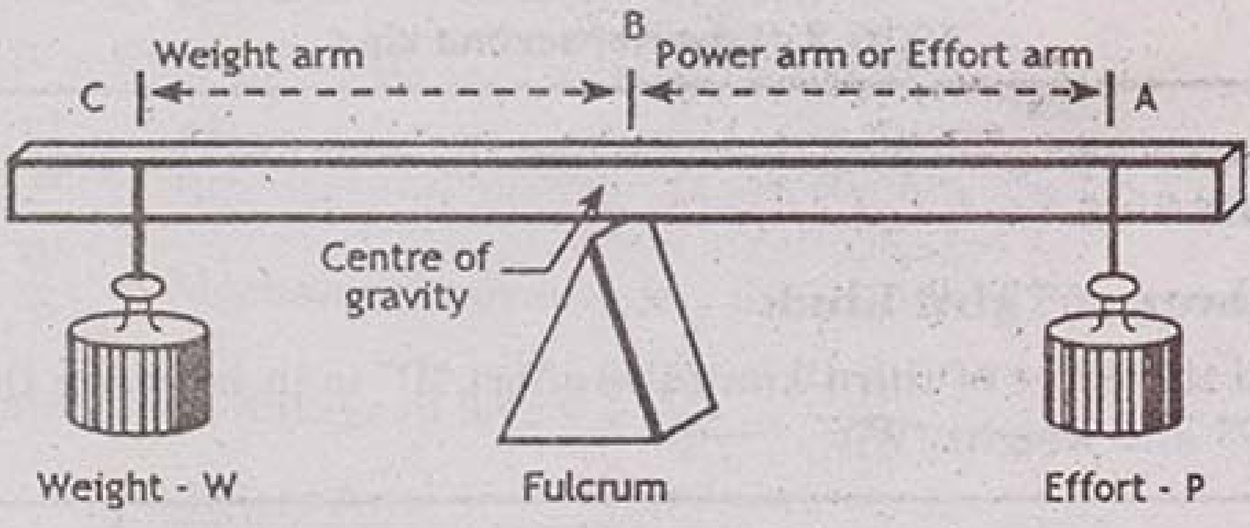


Fig. 9.1: The Lever

KINDS OF LEVER

There are three kinds of levers depending on the position of the effort, weight and fulcrum.

(a) Lever of First Kind

In the lever of first kind, the fulcrum F lies between load "W" and effort "P".

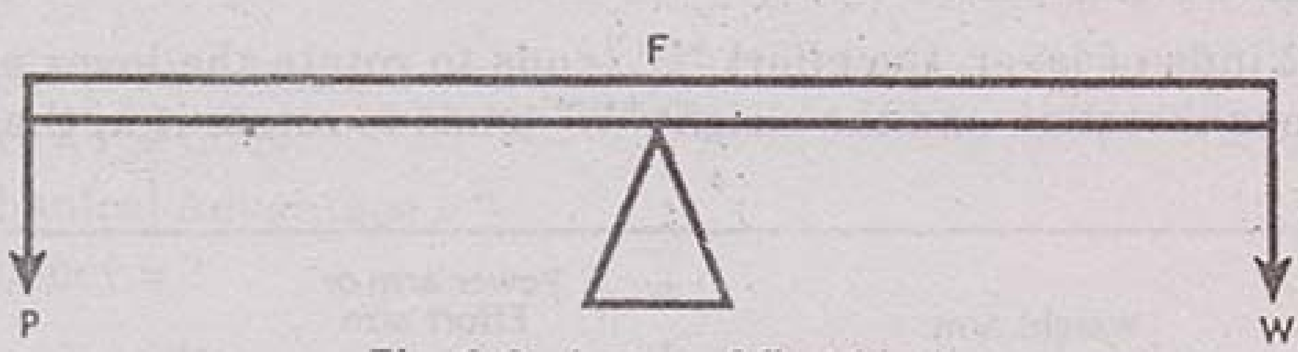


Fig. 9.2: Lever of first kind

An arm balance, a pair of scissors and seesaw are examples of the lever of first kind.

(b) Lever of Second Kind

In the lever of second kind, the weight "W" is in between the fulcrum "F" and the effort "P".

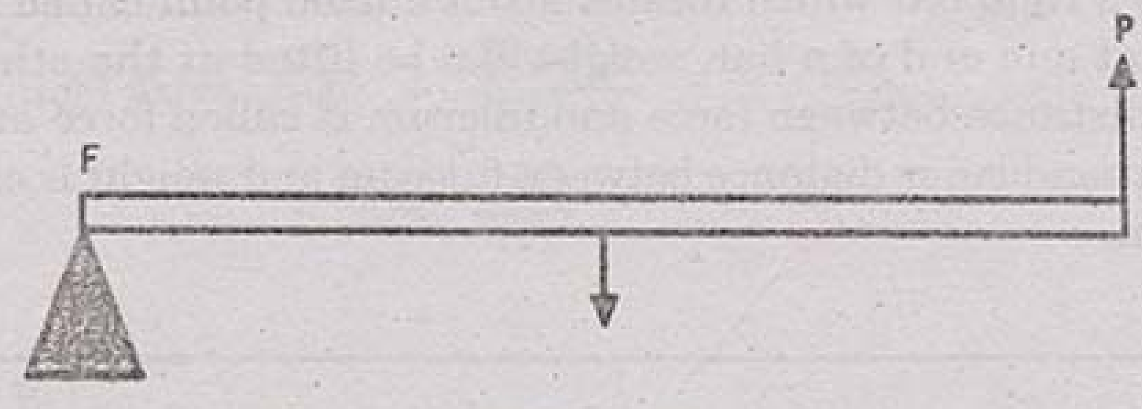


Fig. 9.3: Lever of second kind

The nutcracker, punching machine and the door are the examples of the lever of second kind.

(c) Lever of Third Kind:

In the lever of third kind, the effort "P" is in between the fulcrum "F" and weight "W".

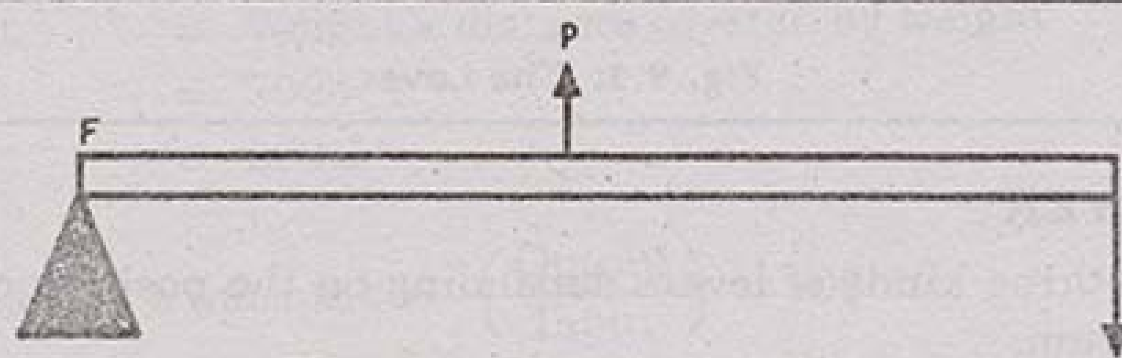


Fig. 9.4: Lever of third kind

A pair of forceps, human arm, a fire tongs are the examples of third kind of lever.

PRINCIPLE OF LEVER:

In all kinds of lever, the effort "P" tends to rotate the lever about the fulcrum in one direction and the weight "W" tends to rotate it in the opposite direction.

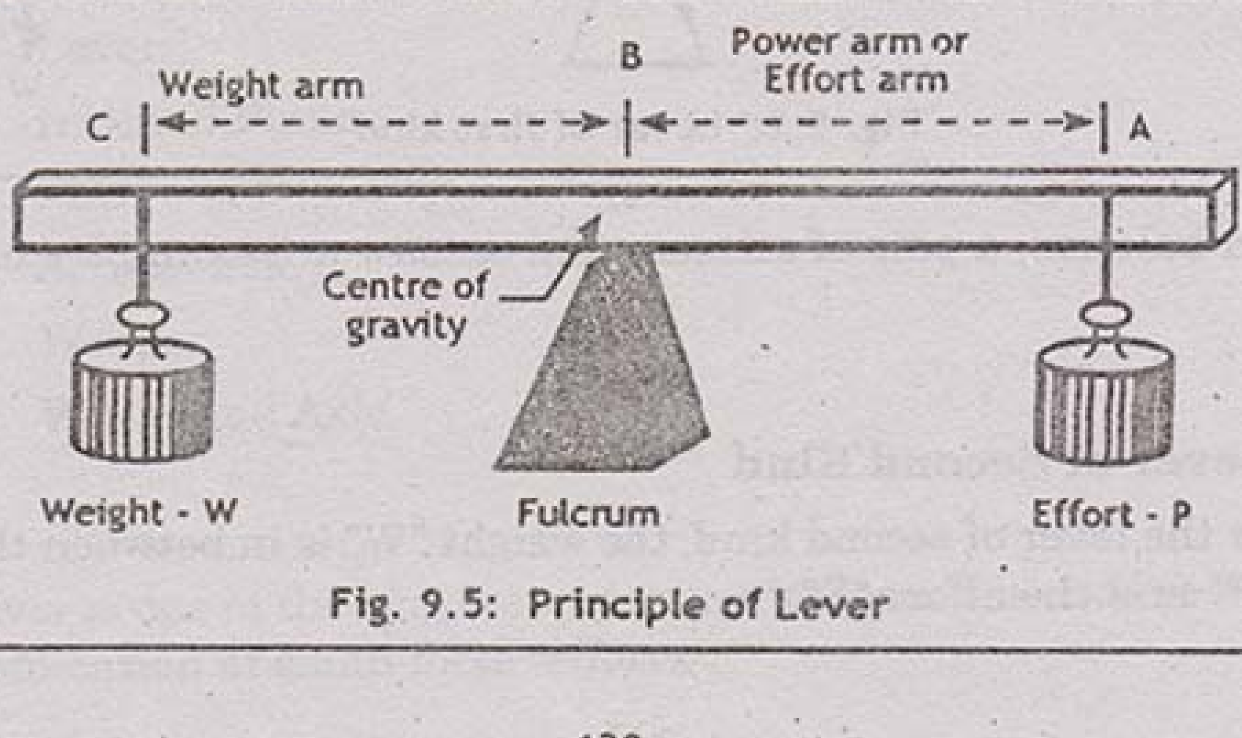


Fig. 9.5: Principle of Lever

The turning effect of any force is called torque, which is equal to the product of force and perpendicular distance.

If the two torques are equal then the lever is said to be in equilibrium. Therefore from fig 9.5, we have

Torque of the effort = Torque of the weight

Weight × weight arm = effort × effort arm

or $W \times BC = P \times AB$

or $\frac{W}{P} = \frac{AB}{BC}$

But $\frac{W}{P}$ = Mechanical advantage

Mechanical advantage of lever $= \frac{W}{P} = \frac{AB}{BC}$

$$= \frac{\text{Effort Arm}}{\text{Weight Arm}} \qquad (9.6)$$

Thus the mechanical advantage of a lever can be increased by increasing the length of the effort arm to a maximum possible limit and by decreasing the length of weight arm to a maximum possible limit.

Example 9.3

A lever overcomes a resistance of 400 N when an effort of 50 N is applied on it. Calculate the mechanical advantage and the efficiency of lever if the effort arm is 10 times as long as weight arm.

Solution:

Weight or resistance = W = 400 N.
Effort = P = 50 N.
Mechanical Advantage = ?
Efficiency = ?

Suppose, "X" meters be the length of the weight arm.
Length of the effort arm = h = 10 X.

Mechanical Advantage $= \frac{\text{Weight}}{\text{Effort}}$

$$= \frac{400}{50} = 8$$

$$\text{Output} = \text{Weight} \times \text{Weight Arm}$$
$$= 400 \times X$$
$$\text{Input} = \text{Effort} \times \text{Effort Arm}$$
$$= 50 \times 10\,X$$
$$\text{Efficiency} = \left(\frac{\text{Output}}{\text{Input}}\right) \times 100$$
$$= \left(\frac{\cancel{400} \times \cancel{X}}{\cancel{50} \times 10\cancel{X}}\right) \times \cancel{100}$$
$$= 80\%$$

THE INCLINED PLANE:

It is our daily experience that it is easier to roll up a heavy object by means of a sloping plank than to lift it vertically to the same height. The sloping plank is an example of an inclined plane.

An inclined plane is actually a rigid plane which is kept inclined to the horizontal at a certain angle.

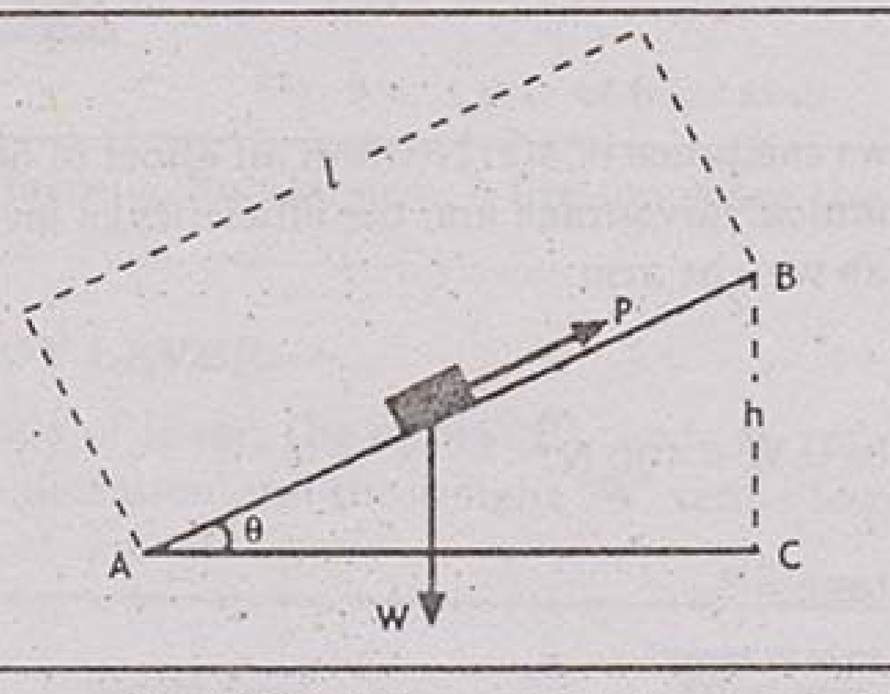


Fig. 9.6:
An Inclined Plane

Fig. 9.6 shows a load "W" being pulled along an inclined plane AB, inclined at an angle θ to the horizontal.

The weight of the load acts vertically downwards. In order to raise it to a vertical height h, the effort "P" has to move distance "l" along the path AB.

$$\text{Input} = \text{Work done by effort}$$
$$= \text{Effort} \times \text{distance}$$
$$= P \times l$$
$$\text{Output} = \text{Work done in raising the load}$$
$$= \text{Weight} \times \text{Height}$$
$$= W \times h$$

For an ideal machine

$$\text{Input} = \text{Output}$$
$$P \times l = W \times h$$

or $\dfrac{l}{h} = \dfrac{W}{P}$

or $\dfrac{W}{P} = \dfrac{l}{h}$

$\dfrac{W}{P}$ = Mechanical Advantage

Mechanical Advantage of an inclined plane $= \text{M.A} = \dfrac{W}{P} = \dfrac{l}{h}$

but $\dfrac{l}{h} = \dfrac{1}{\text{Sin}\,\theta}$

$\therefore$ Mechanical Advantage of an inclined plane $= \dfrac{1}{\text{Sin}\,\theta}$ ——— (9.7)

Equation 9.7 shows that the smaller the value of angle θ, the greater will be the mechanical advantage.

Example 9.4

An inclined plane 5 m long has one end raised by 1 m. Calculate Mechanical Advantage.

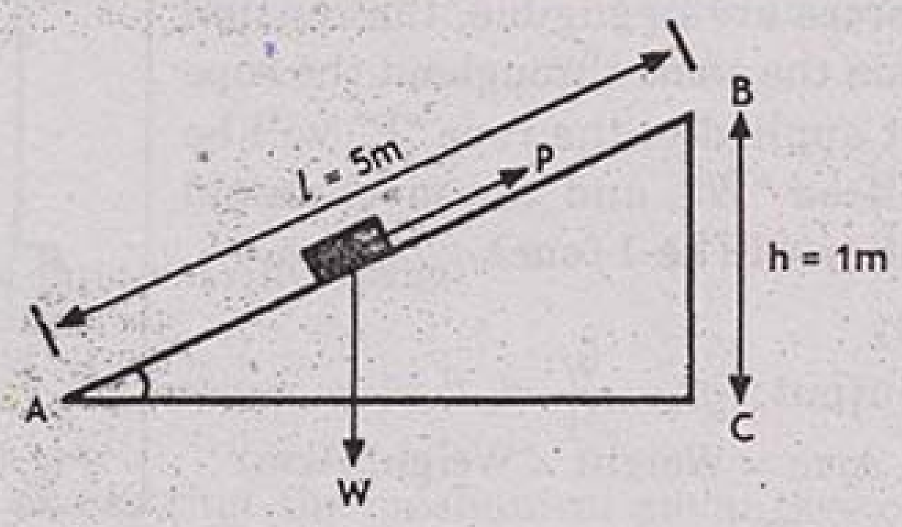


Fig. 9.6(a)

Solution:

Length of the inclined plane = l = 5 m
Height through which the end of the plane is raised = h = 1 m
Mechanical Advantage = M.A = ?

Mechanical Advantage of inclined plane = M.A = $\frac{1}{\sin\theta}$

In fig (9.6-a)

$$\sin\theta = \frac{BC}{AB}$$

$$\sin\theta = \frac{1}{5}$$

$$M.A = \frac{1}{\sin\theta}$$

$$M.A = \frac{1}{1/5}$$

$$M.A = 5$$

PULLEY:

A pulley is a wheel with a grooved rim. The wheel is supported in a frame which is called block. The wheel can turn freely about an axle in the block. It can also be suspended from a fixed beam by means of a hook.

FIXED PULLEY:

A simple pulley is hung on a suitable support with a rope passing around its groove as shown in fig 9.7. A load or weight "W" is tied at one end of the rope passing over a pulley, while the effort is applied downward at the other end. If the weight of the rope and the frictional forces are negligible, the tension in the rope will remain the same throughout the rope and hence the effort applied to the rope "P" will be equal to the load lifted "W", and the mechanical advantage of the pulley will be 1 (one).

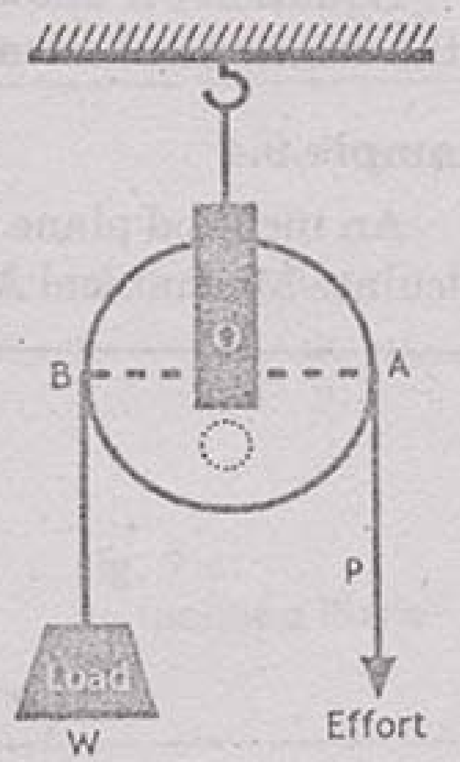


Fig. 9.7: Simple Pulley

Mathematically

Input = Output

Effort × Effort Arm = Weight × Weight Arm

$$P \times OA = W \times OB$$

$$\frac{W}{P} = \frac{OA}{OB}$$

$$\therefore OA = OB$$

$$\therefore \frac{W}{P} = 1 \quad \text{———————} \quad (9.8)$$

The mechanical advantage of fixed pulley is 1.

MOVABLE PULLEY:

In this pulley one end of the rope passing round the pulley is tied to a rigid support O and the effort P is applied at the other end of the rope as shown in Fig 9.8.

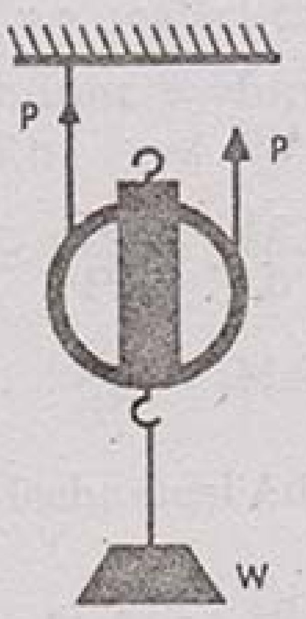


Fig. 9.8

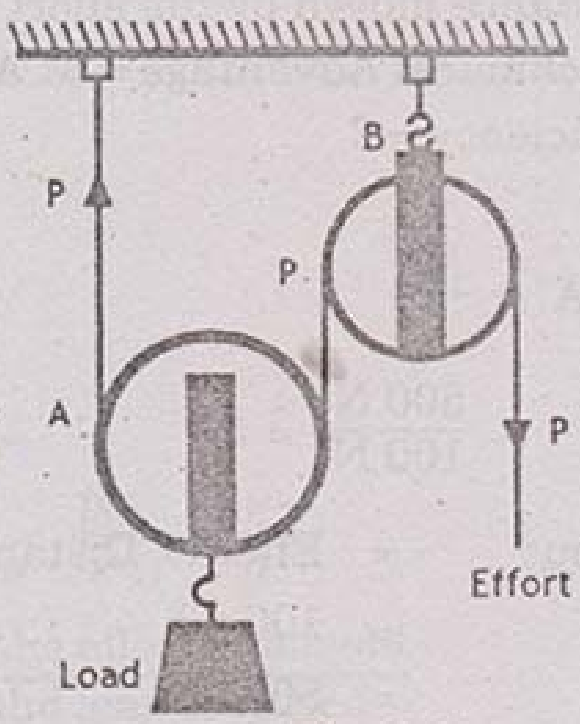


Fig. 9.9

The weight or load W to be lifted is hung from the hook of the block. At every point in the rope, the tension is equal to the applied effort P. As both ends of the rope are pulling the weight "W" upwards so the effort acting on the weight in the upward direction will be 2 P. If we neglect the weight of the rope and friction, then in equilibrium position

Load = 2 x effort

$$W = 2P$$

$$\frac{W}{P} \quad 2$$

$$\therefore \quad \frac{W}{P} = \text{Mechanical Advantage}$$

$$\therefore \quad \text{Mechanical Advantage} = \text{M.A} = \frac{W}{P} = 2$$

Equation 9.9 shows that the mechanical advantage of a movable pulley is 2. It means double load can be lifted with the help of a single movable pulley as compared to effort. The direction of effort can be changed by using another fixed pulley, so that the effort is applied with convenience as shown in fig 9.9.

Example 9.5

A load of 500 N is lifted through 25 cm vertically upward by applying an effort of 100 N upto 5 m on a pulley system. Find the mechanical advantage and the efficiency of the pulley system.

Solution:

Load = W = 500 N

Distance moved by load = h = 25 cm = 0.25 m

Effort = P = 100 N

Distance moved by the effort = d = 5 m

Mechanical Advantage = M.A = ?

Efficiency = ?

$$M.A = \frac{W}{P}$$

$$= \frac{500\text{ N}}{100\text{ N}} = 5$$

$$\text{Input} = \text{Effort} \times \text{Distance moved by the effort}$$

$$= 100 \times 5$$

$$= 500\text{ J}$$

$$\text{Output} = \text{Load} \times \text{Distance moved by the load}$$

$$= 500 \times 0.25 = 125\text{ J}$$

$$\text{Efficiency} = \left(\frac{\text{Output}}{\text{Input}}\right) \times 100$$

$$= \left(\frac{125}{500}\right) \times 100$$

$$= 25\%$$

SCREW JACK:

A screw jack is a simple machine which is commonly used to lift heavy loads such as motor car. It consists of a long screw rod R passing through a threaded block "B" and handle "H" which is called Tommy-bar to turn the threaded block B. As a result it rises up and lifts the heavy load. The distance between two successive screw threads is called pitch. When the effort is applied to the handle H, the effort moves in a circle of radius "r" where r is the length of Tommy bar. In this rotation block B moves up.

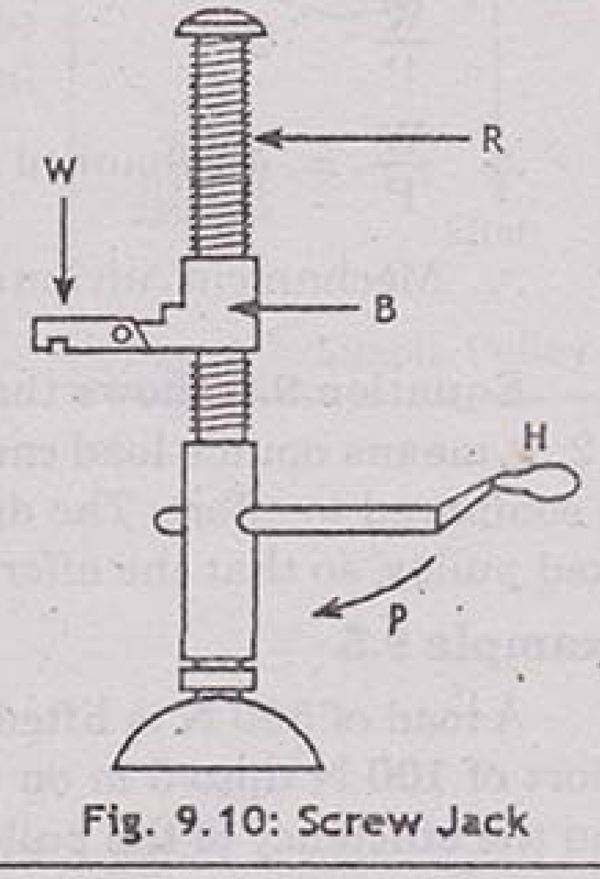


Fig. 9.10: Screw Jack

When the handle is turned through one complete revolution in a circle of radius "r", the effort moves through a distance $2\pi r$ and the load is raised through a height "h" in this case.

$$\text{Input} = P \times 2\pi r$$
$$\text{Output} = W.h.$$

If we neglect the energy losses.

$$\text{Output} = \text{Input}$$
$$W.h. = P \times 2\pi r$$

$$\text{or} \quad \frac{W}{P} = \frac{2\pi r}{h}$$

$$\therefore \quad \frac{W}{P} = \text{Mechanical Advantage}$$

$$\text{Mechanical Advantage of screw jack} = \frac{W}{P} = \frac{2\pi r}{h}$$

$$\text{M.A} = \frac{\text{Distance through which the effort is moved}}{\text{Height through which the load is raised}}$$

As the pitch of the screw is very small as compared to the length of the Tommy-bar so the mechanical advantage of screw jack is very large.

Example 9.6

The length of the Tommy-bar of a screw jack is 25 cm and its pitch is 2.0 mm. How much load can be lifted by applying an effort of 50 N on the Tommy-bar? Also calculate the mechanical advantage of the screw jack.

Solution:

Length of the Tommy-bar = r = 25 cm = 0.25 m

Pitch = h = 2.0 mm = 0.002 m

Effort = P = 50 N

Load = W = ?

Mechanical Advantage = M.A = ?

Since $\quad \text{Input} = \text{Output}$

$$W \times h = P \times 2\pi r$$

$$\frac{W}{P} = \frac{2\pi r}{h}$$

$$W = \frac{2\pi r}{h} P$$

$$= \frac{2 \times 3.142 \times 0.25 \times 50}{0.002}$$

$$= 39275 \text{ N}$$

$$\text{Mechanical Advantage} = \text{M.A} = \frac{2\pi r}{h}$$

$$= \frac{2 \times 3.142 \times 0.25}{0.002}$$

$$= 785.5$$

THE WHEEL AND AXLE:

It is a very simple machine in use for many centuries for lifting heavy loads such as pulling a bucket of water from a deep well. The wheel and axle consist of two cylinders one of larger radius "R" and the other of smaller radius "r" which are mounted on the same shaft which is a common axis of rotation. The cylinder with larger radius is called the wheel while the cylinder with smaller radius is called the axle. The shaft is gripped in clamps so that it can be rotated freely.

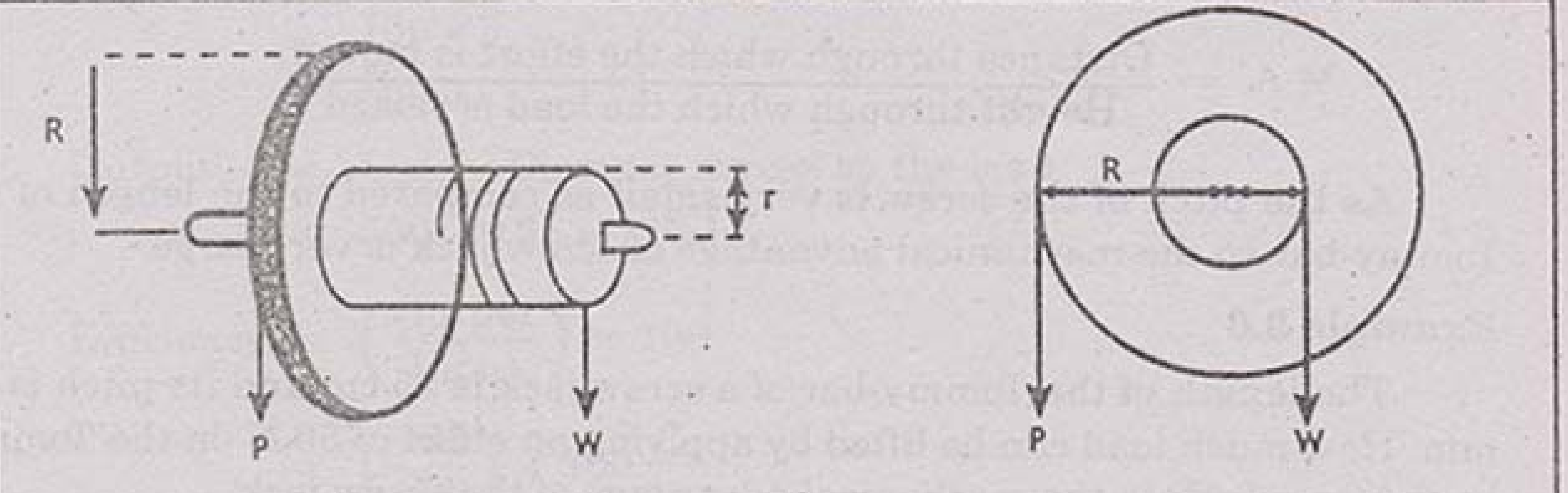


Fig. 9.11: The Wheel and Axle

The effort is applied at the end of a rope wound round the wheel and the load is tied to a rope wound round the axle in the opposite direction. If the rope round the wheel is unwound, it will cause more rope to be wound around the axle and so it lifts the load.

When the effort applied turns the wheel through one complete revolution, the axle also turns through one revolution. Thus in the same interval of time, the effort will move through a distance of $2\pi R$ and the load will move through a distance $2\pi r$. If the forces of friction are negligible then,

$$\text{Output} = \text{Input}$$

$$W \times 2\pi r = P \times 2\pi R$$

$$\frac{W}{P} = \frac{\pi R}{\pi r}$$

Thus mechanical advantage (M.A) of the wheel and axle system is

$$\text{M.A.} = \frac{R}{r}$$

SUMMARY

1. **Simple Machine:** Any thing which helps man in doing work more easily is called simple machine. Lever, pulley, inclined plane, screw jack and wheel and axle are simple machines.
2. **Effort:** The force applied on the machine is called effort.
3. **Load:** Weight lifted or the resistance which is over come by a machine is called load.
4. **Mechanical Advantage:** Mechanical Advantage $= \frac{\text{Load}}{\text{Effort}}$
5. **Input:** It is the work done on the machine.
6. **Output:** It is work done by the machine.
7. **Efficiency:** The ratio of the useful work done by a machine to the work done on its is called efficiency.
8. **Lever:** A strong wooden or metallic bar, which can be rotated about a point, is called a lever.
9. **Fulcrum:** The point around which lever is rotated.
10. **Effort Arm:** The straight distance between fulcrum and effort.
11. **Load Arm:** The straight distance between fulcrum and load.
12. **Moment of Effort:** The product of effort and effort arm is called moment of effort.
13. **Moment of Load:** The product of load and load arm is called moment of load.
14. **Principle of Lever:** Effort x Effort Arm = Load x Load Arm.
15. **Pulley:** Pulley is a wheel type disc. A slot is made on its circumference. It is supported by a block.
16. **Inclined Plane:** A surface whose one end is higher than the other end is called an inclined plane.

$$\text{M.A.} = \frac{L}{E} = \frac{l}{h}, \text{ or M.A.} = \frac{1}{Sin\ \theta}$$

17. **Screw Jack:** It is a simple machine which consists of a nut and bolt through which heavy load is lifted easily. For screw jack:

$$\text{M.A} = \frac{2\pi l}{h}$$

18. **Wheel and Axle:** It is a simple machine containing two cylinders of different diameters, connected with a common shaft used for lifting heavy load. For wheel and axle:

$$\text{M.A.} = \frac{R}{r}$$

QUESTIONS

9.1 Write answer to the question given below.

(i) What is a lever? Determine its mechanical advantage.

(ii) To what kind of lever the following machines belong:
Door, humanarm, forceps, see-saw, pair of scissors, hand cart, balance, handle of hand pump, upper and lower jaws of mouth.

(iii) What is an inclined plane and how does it help in doing work?

(iv) Determine the mechanical advantage of an inclined plane.

(v) Describe the construction of a simple pulley.

(vi) What type of work is done by a fixed pulley and a movable pulley.

(vii) In what ways are a lever, an inclined plane and a pulley are alike

9.2 Fill in the blanks.

(i) When we assume that there is no loss of ________ due to friction in a machine then the ________ is equal to the ________.

(ii) The efficiency of a machine is the ________ of work done by the machine to the work done ________.

(iii) The see-saw is an example of a order of ________ kind.

(iv) The mechanical advantages of a wheel and axle system is given by the ________ of the ________ to the ________.

(v) ________ is a device which makes the process of doing work easier.

9.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) A ________ is a simple machine.
(a) first (b) second (c) third

(ii) If the fulcrum of a lever is between the effort and resistance, it is a ________ class lever.
(a) first (b) second (c) third

(iii) A pair of scissors is an example of a ________.
(a) pulley (b) lever (c) wheel and axle (d) inclined plane

(iv) Which of the following belong to the second kind of lever.
(a) pair of scissor (b) pair of forceps (c) door

(v) A screw jack always has an efficiency less than 100%. Which one of the following is the best explanation of the reason for this?
(a) this must be so because of the definition of screw jack
(b) a screw is really an inclined plane

(c) there is always a frictional force between its parts
(d) a screw jack only acts vertically and hence has weight
(e) a screw jack is not a totally efficient machine.

9.4 Pick out true and false from the following sentences.

(i) For a screw jack the mechanical advantage is given by $2\pi d/h$, where d is the effort arm and h is the pitch of the screw.

(ii) The efficiency of an inclined plane is increased when the angle it makes with the horizontal is decreased.

(iii) The mechanical advantage of a movable pulley is four.

(iv) The ratio of output and input in a machine is known as Mechanical advantage.

(v) In a fixed pulley its block is connected from the weight.

PROBLEMS

9.1 An effort of 160 N, applied to a pulley system is able to lift a load of 750N through a vertical height of 1.2 m. To do this, the effort moves a distance of 7.2 m. Calculate (a) the mechanical advantage of the pulley system (b) the efficiency.

(4.7, 78.3%)

9.2 The ratio between the length of the arms of a lever is 4:1. How much force be applied to the longer arm so that a load of 20 N suspended at the shorter arm can be balanced?

(5 N)

9.3 An object of mass 100 Kg is raised 2m above the ground using an inclined plane of length 10m. Calculate the effort applied parallel to the inclined plane.

(196 N)

9.4 An effort of 100 N, applied to a pulley system, is able to lift a load of 400 N through a vertical height of 1 m. To do this, the effort moves a distance of 8 m. Calculate (a) the mechanical advantage of the pulley system, (b) the efficiency.

(4, 50%)

9.5 A load of 21000 N placed on the top of a screw jack is lifted by a force of 300 N. If the length of the Tommy-bar is 35 cm and the pitch of the screw is 2.5 mm, find its mechanical advantage and efficiency.

(70, 79.5%)

9.6 The length of the handle of a screw jack is 40 cm and its pitch is 4 mm How much load can be lifted by applying a force of 5 N on the handle.

(3142.86 N)

CHAPTER - 10

PROPERTIES OF MATTER

LEARNING OBJECTIVES:

- Elasticity.
- Comparison of elasticities of two substances.
- Hooke's law applied to a helical spring.
- Pressure.
- Atmospheric pressure.
- Demonstration of atmospheric pressure.
- Pressure of liquids and gases.
- Pascal's law.
- Applications of Pascal law.
- Archimedes principle.
- Analytical treatment of Archimedes principle.
- Buoyancy and law of floation.
- Applications of Archimedes principle.
- Kinetic molecular theory of matter.
- Surface tension.
- Examples showing surface tension.
- Fluid friction (viscosity).

INTRODUCTION

The most fundamental property of matter which is directly concerned with its very existence is that it has mass and occupies space irrespective of its state. All the other properties are the derived properties like density, pressure of a gas, surface-tension and viscosity of liquids etc.

In general matter exists in three states or phases. They are solid, liquid and gas.

The solid bodies (by which we always mean rigid bodies) have fixed volume and fixed shape. In solids the molecules are closely packed. The force of attraction between them is very strong. Actually, it is because of the strong force of attraction that the solid bodies have fixed shape and fixed size. The motion of molecules in a solid body is very much restricted. They just vibrate about their mean position. They do not leave their mean position. The relative distance between the atoms or molecules of a solid body is fixed.

Liquids have fixed volume but no fixed shape. They assume the shape of the container. The force of attraction between the molecules is small as compared to that found in solids. The liquid molecules move freely inside the volume occupied by the liquid. Liquids are incompressible and maintain their level.

Gases have neither fixed volume nor fixed shape. They can occupy any space available to them. The force of attraction between the molecules of a gas is negligibly small. This is the reason why the gas molecules move freely in all directions with all possible velocities.

In general the state of matter depends on its temperature. No matter is found in the same state at all temperatures. We explain this statement by taking a simple and common example of water. At an ordinary room temperature (in our country), water exists as a liquid. If the water is cooled down to 0°C it becomes solid which we call ice. However, if water at room temperature is heated to 100°C at atmospheric pressure, it is vapourized and changes into steam which is the gaseous state of water.

Matter is also found in a fourth state called plasma. This is an ionized state of matter. In the most general sense, a plasma is any state of matter which contains enough free charged particles for its dynamical behaviour to be governed by electromagnetic waves. Thus it encompasses the solids since electrons in metals and semiconductors fall into this category; however, plasma is largely concerned with ionized gases.

10.1 ELASTICITY

There are particular substances which undergo the change in shape or size. Such change is called deformation. When a force is applied to them and regain their original state after the removal of the force, such bodies are called elastic bodies and this property possessed by them is called elasticity. For example, a force is applied to stretch a rubber band which regains its original sate after the removal of the force. Thus a rubber band is an elastic body. Similarly a wooden metre scale bends when a force is applied to it. After the removal of the force it straightens out. Hence such a metre scale is an elastic body.

If a body is deformed and is kept in the state of equilibrium. The apparent force acting on a deformed body is the external applied force. To keep the body in equilibrium the body must be acted upon by some force other than the external deforming force. This internal force arising from the deformation of the body and acting per unit area of the body is called **stress.** The relative deformation produced in the body is called **strain.** For an elastic body the ratio of stress to strain is a constant and is called the modulus of elasticity. This relation was

first discovered by Robert Hooke and is known as Hooke's law – which states that for any elastic body stress is proportional to strain which we write as:

$$\text{Stress} \ \alpha \ \text{Strain}$$

$$\frac{\text{Stress}}{\text{Strain}} = \text{Constant} \text{ ——————— (Hooke's Law)} \quad (10.1)$$

As mentioned above stress is defined as the magnitude of the force normal to the surface acting per unit area of the body. If F stands for force and A for area, then the stress is given as:

$$\text{Stress} = \frac{F}{A} \text{ ——————————————— } (10.2)$$

As we know strain is the relative deformation. The deformation may be the change in length, change in volume or change in shape. We limit ourselves to the change in length. If L is the original length and ΔL is the change in length of an elastic wire under the deforming force.

Then the longitudinal strain is given as:

$$\text{Longitudinal strain} = \frac{\Delta L}{L} \text{ ——————————— } (10.3)$$

The ratio of the stress to the longitudinal strain is called Young's modulus of elasticity. It is denoted by Y and is given as

$$Y = \frac{\text{stress}}{\text{longitudinal strain}}$$

$$Y = \frac{F/A}{\Delta L/L} = \frac{F \times L}{A \times \Delta L} \text{ ——————————— } (10.4)$$

It is clear from eq.(10.4) that the unit of Young's modulus of elasticity (in the system international) is N/m^2 which is also the unit of stress.

Hooke's law is obeyed upto a certain limit called the elastic limit, that is, there is certain maximum value of the stress which deforms a body beyond which the body will not regain its original state after the removal of the force (or the stress). This means a body retains its elasticity upto a certain limit and Hooke's law is obeyed for a certain maximum value of the stress. If the stress exceeds this maximum value, then the body will be permanently deformed. It cannot return back to its original state. The body is said to have crossed the elastic limit. Here we give a list of the values of Young's modulus of elasticity and the elastic limit for some materials.

Table 10.1

Values of Young's modulus and elastic limits for some materials

Material	Young's Modulus (Nm^{-2})	Elastic limit (Nm^{-2})
Aluminium	7.0×10^{10}	1.3×10^{8}
Brass (cast)	13.1×10^{10}	3.8×10^{8}
Copper	11.5×10^{10}	8.0×10^{8}
Iron	19.4×10^{10}	7.7×10^{8}
Lead	1.6×10^{10}	——
Steel	20.0×10^{10}	2.7×10^{8}
Rubber (vulcanised)	14.0×10^{4}	——
Tungsten	36.0×10^{10}	——
Crown glass	7.1×10^{10}	1.5×10^{9}

We now define a few terms in connection with elasticity.

Elastic limit:

It is the maximum value of stress within which the body exhibits the property of elasticity. Below the elastic limit, the body regains its original state of shape and size after the removal of the deforming force. Beyond the elastic limit the body does not regain completely its original state when the deforming force is removed.

Example 10.1

The area of cross section of a metallic bar is 1cm by 1cm. Its length is 50cm. A force of 10000N stretches it by 0.25cm. Find the stress and the strain.

Solution:

$L = 50cm = 0.5m$ = original length

$F = 10000N$ = force applied

$A = 1cm \times 1\ cm = \frac{1}{100} m \times \frac{1}{100} m = 10^{-4} m^2$ = area of cross section.

$\Delta L = 0.25\ cm = 0.25 \times 10^{-2} m = 2.5 \times 10^{-3} m$ = change in length

(i) Stress = ?

$$\text{Stress} = \frac{\text{Force}}{\text{Area}} = \frac{10000\ \text{N}}{10^{-4}\text{m}^2} = 1 \times 10^{8}\text{N/m}^2$$

(ii) Strain = ?

$$\text{Strain} = \frac{\text{Change in length}}{\text{Original length}} = \frac{2.5 \times 10^{-3}\text{m}}{0.5\text{m}} = 5.0 \times 10^{-3}$$

Example 10.2

A metal bar of cross sectional area 1000mm^2 is 4m long. A force of 5000N increases its length by 0.25mm. Find Young's modulus of the metal.

Solution:

F = 5000 N = applied force

A = $1000\ \text{mm}^2 = 1000 \times 10^{-3}\text{m} \times 10^{-3}\text{m} = 10^{-3}\text{m}^2$ = area of cross section

L = 4m = original length

ΔL = $0.25\text{mm} = 0.25 \times 10^{-3}\text{m} = 2.5 \times 10^{-4}\text{m}$ = change in length.

Y = ? = young's modulus

Formula used

$$Y = \frac{F \times L}{A \times \Delta L} = \frac{5000\text{N} \times 4\text{m}}{10^{-3}\text{m}^2 \times 2.5 \times 10^{-4}\text{m}}$$

$$Y = 1.6 \times 5 \times 10^{10}\ \text{N/m}^2$$

$$Y = 8.0 \times 10^{10}\ \text{N/m}^2$$

Example 10.3

A steel wire 4m long has an area of cross section equal to $2.4 \times 10^{-7}\text{m}^2$. It is stretched by a force of 36N. Calculate (i) stress, (ii) strain, (iii) increase in length. Given $Y = 1.8 \times 10^{11}\text{N/m}^2$.

Solution:

$$L = 4m = \text{length of the wire}$$
$$A = 2.4 \times 10^{-7}m^2 = \text{area of cross section}$$
$$F = 36N = \text{applied force}$$
$$Y = 1.8 \times 10^{11}N/m^2$$

(i) Stress = ?

$$\text{Stress} = \frac{F}{A} = \frac{36N}{2.4 \times 10^{-7}m^2} = 1.5 \times 10^{8}N/m^2$$

(ii) Strain = ?

$$\therefore Y = \frac{\text{Stress}}{\text{Strain}}$$

$$\text{Strain} = \frac{\text{Stress}}{Y} = \frac{1.5 \times 10^{8}N/m^2}{1.8 \times 10^{11}N/m^2}$$

$$\text{Strain} = 8.33 \times 10^{-4}$$

(iii) ΔL = ? = Increase in length

$$\Delta L = L \times \text{strain}$$
$$= 4m \times 8.33 \times 10^{-4} = 33.32 \times 10^{-4}m$$
$$= 3.3 \text{ mm}$$

10.2 COMPARISON OF ELASTICITIES OF TWO SUBSTANCES

It is our observation that some elastic substances undergo deformation by applying a relatively small stress where as some other substances need large values of stress to bring about the same strain. This means that some substances are more elastic than others.

A body is said to be more elastic if it retains its elasticity on the application of a large external force than to a body which loses its elasticity at a lesser external force.

To make it clear, consider two elastic substances x and y in the form of wires. They are geometrically identical, that is, they have the same length and same area of cross section. Let us suppose that the same stress is applied to both of them. We further suppose that $\in_x$ and $\in_y$ are the strains produced in x and y respectively. Let E_x and E_y be the modulii of elasticities of the two bodies respectively.

By definition

$$Ex = \frac{\sigma}{\epsilon_x} \quad \text{and} \quad Ey = \frac{\sigma}{\epsilon_y}$$

Where

σ = stress applied

$$\therefore \quad \frac{Ex}{Ey} = \frac{\sigma/\epsilon_x}{\sigma/\epsilon_y} = \frac{\epsilon_y}{\epsilon_x}$$

If $\epsilon_x > \epsilon_y$, then

$Ex < Ey$

Thus we conclude that a body with smaller strain is more elastic than the body having a larger strain when the applied stress same. Thus we can easily say that steel is more elastic than rubber. In simple words we can say that a body which is easily deformed is less elastic than the body which undergoes deformation with difficulty when the applied stress is same.

10.3 HOOKE'S LAW APPLIED TO A HELICAL SPRING

Let a helical spring be suspended vertically from a fixed support. A block is attached at its lower end. Due to force of gravity the block is displaced downward. Fig (10.1). As a result the spring is stretched. After a little while the block comes to rest and attains equilibrium. Under this condition the block is under the action of two forces (i) the force of gravity acting down ward and (ii) the tension in the spring in upward direction. In equilibrium the two forces are equal and opposite.

According to Hooke's law;

"Tension is proportional to extension".

If F stands for tension and x for extension, then

$$F \propto x$$

$$F = -Kx \quad \text{————————} \quad (10.5)$$

Where 'K' is a constant of proportionality and is known as the spring constant. Its unit is N/m. In the above equation negative sign shows that tension and extension act in opposite directions.

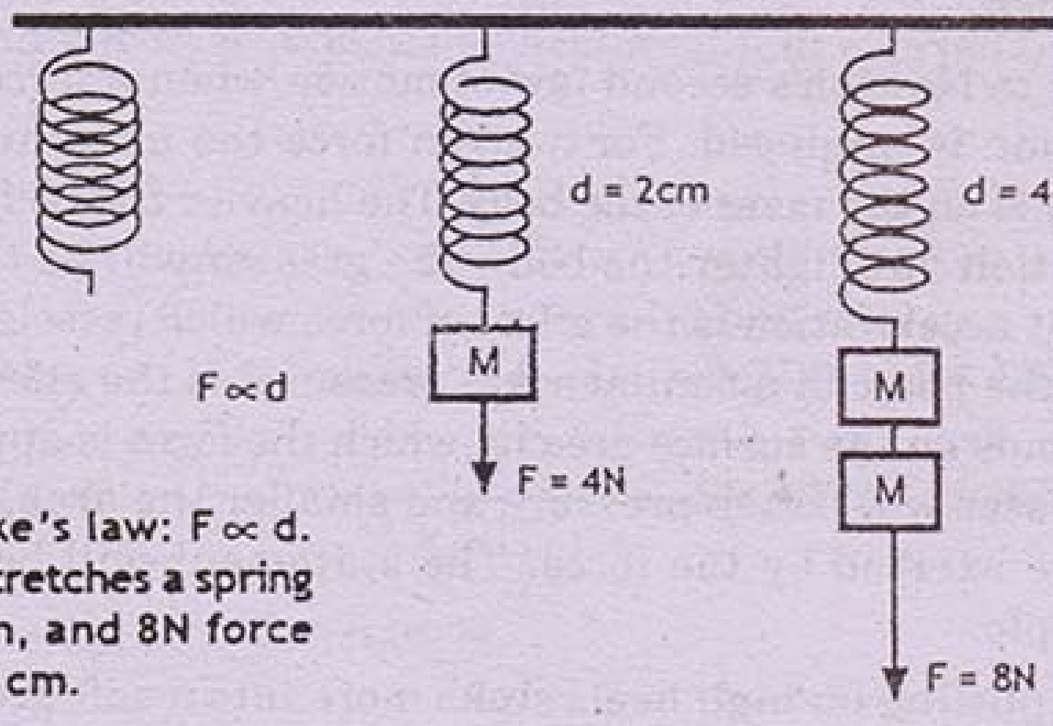


Fig. 10.1: Hooke's law: $F \propto d$. If a force of 4N stretches a spring a length of 2 cm, and 8N force will stretch it 4 cm.

Example 10.4

A force of 10N produces an extension of 2.5Cm. (a) What will be the extension if a force of 25N is applied to the spring, assuming that the elastic limit is not crossed. (b) What is the force required to stretch the spring by 4 Cm?

Solution:

(a) $F_1 = 10N$ = force applied

$x_1 = 2.5Cm$ = extension

$F_2 = 25N$

$x_2 = ?$

From the given values of F_1 and x_1 we calculate the spring constant K as shown below, which is used to find the extension x_2 for the force F_2.

$$F_1 = Kx \Rightarrow F_1 = Kx_1 \quad \text{(only considering magnitude)}$$

$$K = \frac{F_1}{x_1} = \frac{10N}{2.5Cm} = 4N/Cm$$

$$F_2 = Kx_2 \Rightarrow x_2 = \frac{F_2}{K} = \frac{25N}{4N/Cm} = 6.25\ Cm$$

$$x_2 = 6.25\ Cm$$

(b) $x = 4\ Cm$

$K = 4N/Cm$

$F = ?$

$$F = Kx \Rightarrow F = 4N/Cm \times 4Cm = 16N$$

$$F = 16N$$

10.4 PRESSURE

According to Newton's second law of motion when a force is applied to a body, acceleration is produced. For a given force the amount of acceleration produced depends on the mass of the body. The heavier the body, the lesser will be the acceleration and lighter the body, the greater will be the acceleration. This means that acceleration is the effect of force which (acceleration) depends on the mass of the body. In a similar way "pressure" is the effect of force which (pressure) depends on the surface area on which the force is applied. The larger the area, the lesser will be the pressure and smaller the area, the greater will be the pressure exerted by the force. The statement will be clear from the following example.

A woman putting on high heels sinks more into a soft ground than when the same woman puts on flat shoes.

In both the cases the force pulling down is the force of gravity (weight of the woman) acting on the woman. The force of gravity is the same at a place, so why the woman goes deeper in the first case (when wearing high heels) than when she wears flat shoes.

Explanation: No doubt, the gravitational pull is the same in both the cases, but the pressure exerted by the force on the heels (having lesser area) is greater than that exerted on the flat shoes which have large areas.

Definition of pressure:

It is the force acting normally per unit area of a surface. If F stands for force and A for surface area, then the pressure P exerted by the force is given as:

$$\text{Pressure} = \frac{\text{Force}}{\text{Area}}$$

$$P = \frac{F}{A} \qquad \text{———————————— (10.6)}$$

In the system international the unit of pressure is pascal (Pa) which is the same as N/m^2. If one newton force acts normally on a unit area, then the pressure exerted on that area is said to be of one pascal.

Hence

$1 \text{ Pa} = 1 N/m^2$
$2 \text{ Pa} = 2 N/m^2$
and so on

High pressures are expressed in Kilo pascal so that:

1 Kilo pascal (1K Pa) = 1000 pascal

Very surprising result is obtained by comparing the pressure exerted by a woman putting all her weight on to one heel with that exerted by an elephant standing on one foot.

Let us suppose that:

weight of woman = 500 N

$$\text{area of heel} = 1\text{Cm} \times 1\text{Cm} = 0.01\text{m} \times 0.01\text{m} = 1 \times 10^{-4}\text{m}^2$$

$$\text{pressure} = \frac{\text{force}}{\text{area}} = \frac{500\ \text{N}}{1 \times 10^{-4}\text{m}^2} = 5 \times 10^6\text{Pa} = 5000000\ \text{Pa}$$

weight of the elephant = 54000 N

$$\text{area of the foot} = 30\text{Cm} \times 30\text{Cm} \qquad \text{(taking the foot as square)}$$

$$= 0.3\text{m} \times 0.3\text{m}$$

$$\text{pressure} = \frac{\text{force}}{\text{area}} = \frac{54000\ \text{N}}{0.3 \times 0.3} = 600{,}000\ \text{Pa}$$

10.5 ATMOSPHERIC PRESSURE

The atmosphere is the blanket of air around the earth. It extends upto about 1000 Km above the surface of the earth. Since air is a kind of matter, it has weight. The total weight of the atmosphere corresponds to its mass of about 4.5×10^{18} Kg. The pressure exerted by the weight of the atmosphere on the surface of the earth is called "atmospheric pressure". The device used to measure the atmospheric pressure is called barometer. At sea-level its value is about 10^5 pa. This standard pressure is sometimes called one atmosphere or one bar. We do not feel this large pressure because the pressure exerted by blood in our body is nearly the same as the atmospheric pressure.

At high altitudes like mountains, the atmospheric pressure is less which causes difficulty in breathing and nose bleeding may also occur.

10.6 DEMONSTRATION OF ATMOSPHERIC PRESSURE

(1) Collapsing can experiment:

The effect of atmospheric pressure can be demonstrated by evacuating a tin using a vacuum pump. Before the air is pumped out, the pressure inside the tin is equal to the atmospheric pressure. When air is partially removed from the tin by the vacuum pump, the pressure of air inside the tin is less than the atmospheric pressure. Hence the collapsing of the tin takes place as shown in Fig. (10.2).

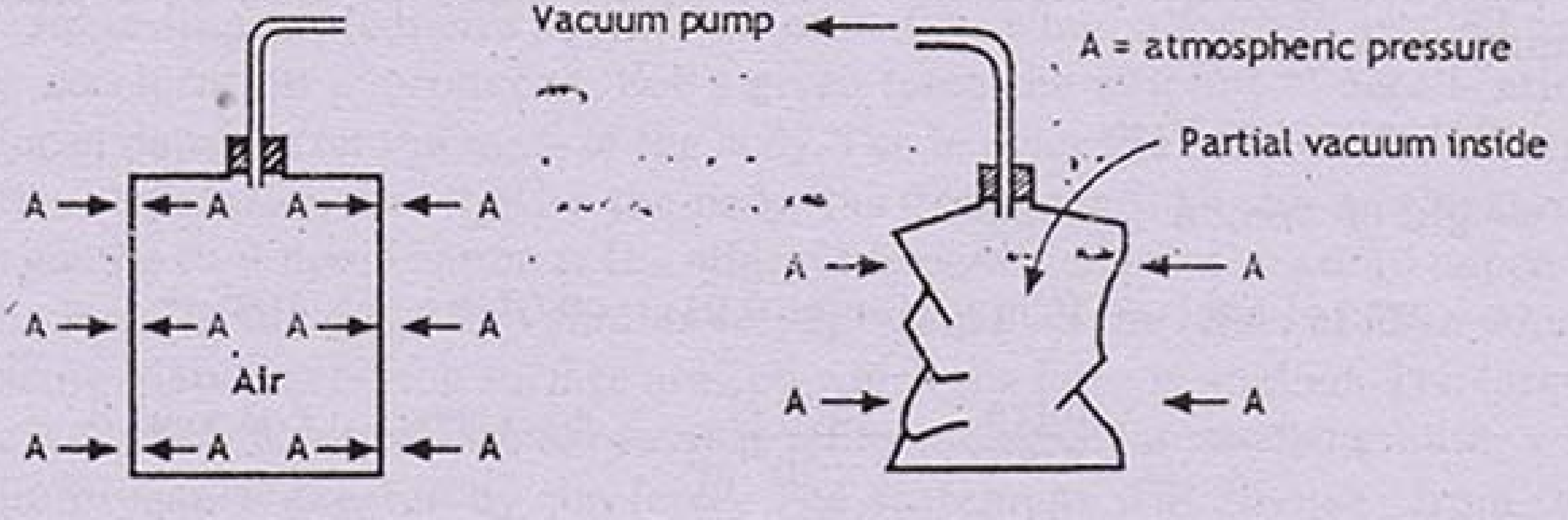


Fig. 10.2: Collapsing of the can after evacuation

(2) Megdeburg-hemisphere experiment:

The very existence of atmospheric pressure was first demonstrated by a German scientist Von Guericke. His experiment is historically known as Magdeburg hemisphere experiment because it was performed first in the city of Magdeburg in Germany. He took two hollow metallic hemispheres. The hemispheres were placed in contact. The air inside the hemispheres was pumped out by a vacuum pump. After the partial removal of the air from inside the hemispheres it was almost impossible to separate them by pulling them apart. It is due to the fact that the pressure exerted by the atmosphere on the outer walls of the hemispheres is much greater than the pressure exerted by the air left inside the hemisphere.

In the original experiment with perfect vacuum inside the hemispheres, even sixteen horses (eight on each side) could not separate the hemispheres. Refer Fig. (10.3)

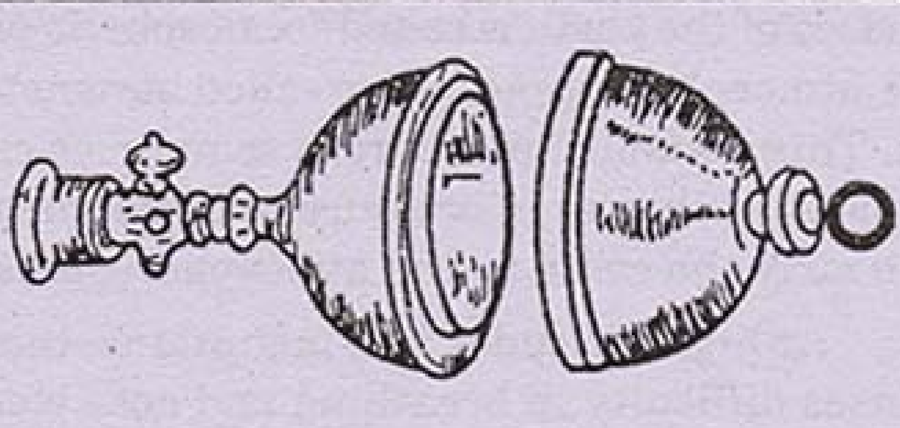


Fig. 10.3: Magdeburg hemisphere apparatus

10.7 PRESSURE OF LIQUIDS AND GASES

Consider a liquid contained in a vessel as shown in the Fig (10.4). Since liquid is a kind of matter, it possesses mass and hence it has weight. It is pulled down vertically by a force equal to its weight and so exerts pressure on the bottom of the vessel. We calculate the pressure of the liquid at a depth 'h' from the free surface of the liquid. For this purpose we imagine a horizontal circle of area 'A' at this depth as shown in the figure. The pressure exerted on this

circle is due to the weight of the liquid contained in a right circular cylinder of height, 'h' and of area of cross section A. We first find the volume of the liquid in this imaginary cylinder, then the corresponding mass of the liquid and finally the weight of the liquid. If V, m and W stand for volume, mass and weight of the liquid respectively, then

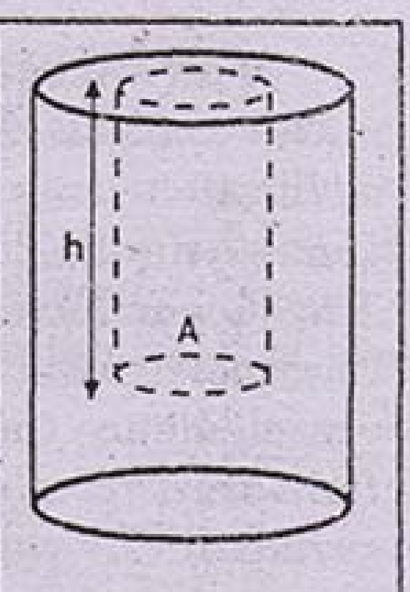


Fig. 10.4

$$V = A \times h$$

The pressure P exerted at any point on the area A is given

$$P = \frac{F}{A} = \frac{W}{A} = \frac{\rho Ahg}{A} = \rho gh$$

$$P = \rho gh \qquad (10.7)$$

The above equation shows that the pressure is directly proportional to the density of the liquid and depth h measured from the free surface of the liquid. For a given liquid the variation of pressure with depth is shown in the Fig (10.5). The pressure of water at spouts A, B and C are shown. They are fitted at different depths h_1, h_2 and h_3 respectively. Here $h_3 > h_2 > h_1$.

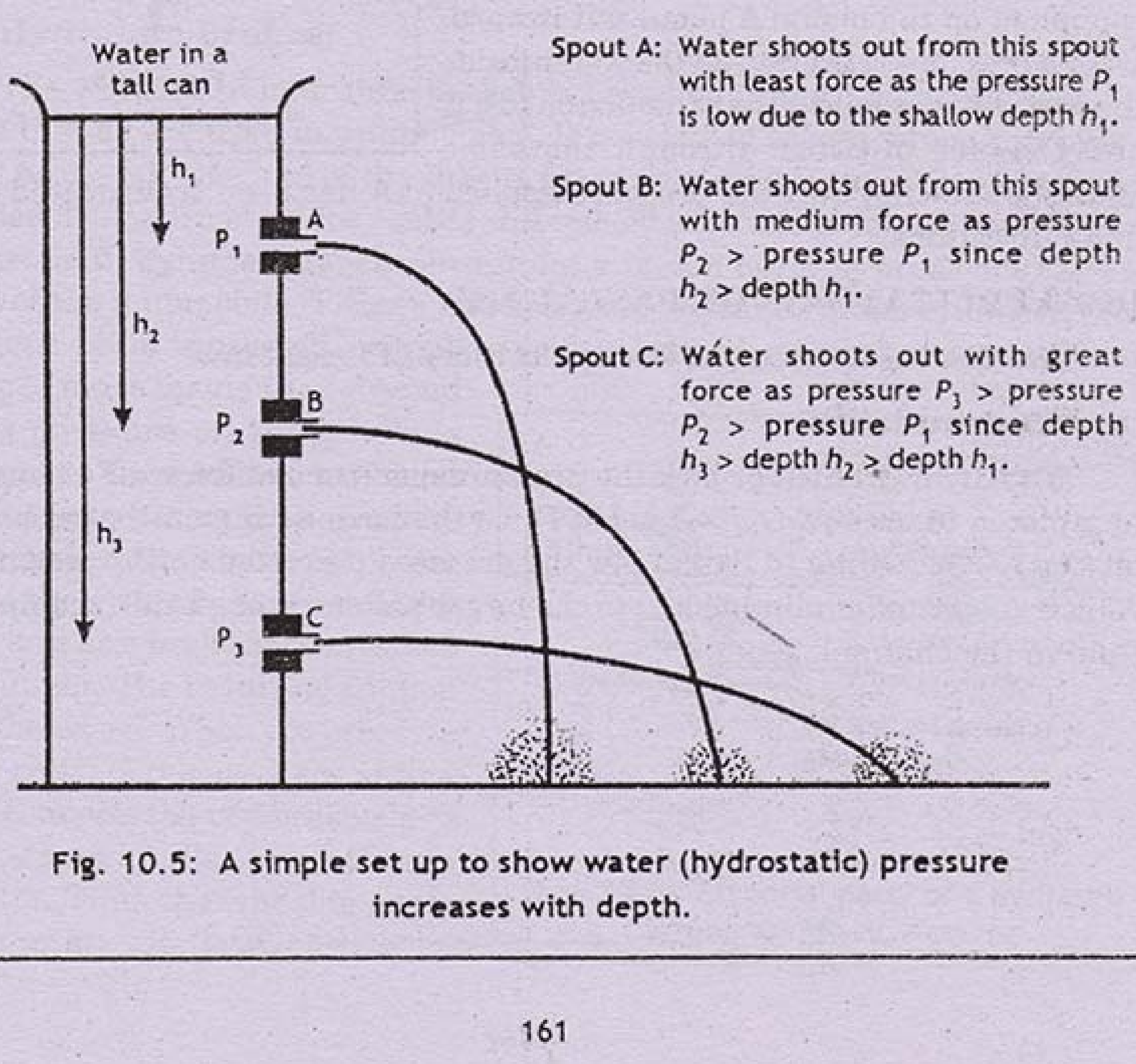


Fig. 10.5: A simple set up to show water (hydrostatic) pressure increases with depth.

Now we describe how gases exert pressure. According to kinetic theory, matter consists of small particles called molecules which are in random motion of vibration, translation or rotation. In gaseous state the molecules are far apart from each other under the ordinary condition of temperature and pressure. They do not have any force of interaction between them except during collision. They move about in all directions with all possible velocities in the containing vessel. Hence they collide constantly with the walls of the vessel. In doing so they exert force on the walls. The force exerted normally per unit area of the surface of the walls is called pressure of the gas.

10.8 PASCAL'S LAW

The law states that when a pressure is applied to a liquid, it is transmitted undiminished equally in all directions. We can prove the statement experimentally.

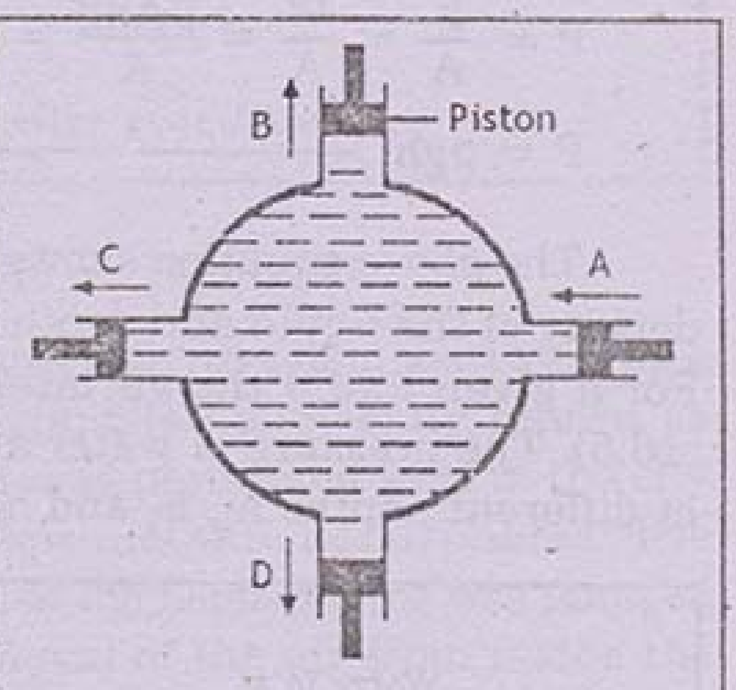


Fig. 10.6:
Pascal Principle

Take a spherical vessel fitted with four water-tight pistons having the same area of cross section as shown in Fig (10.6). The vessel is filled completely with water. Force is applied on the piston A to move it inward. Thus a pressure is exerted on the water inside the vessel. It is found that the pistons B, C and D move outward through the same distance showing that the pressure applied at A has been transmitted equally in all directions.

10.9 APPLICATIONS OF PASCAL LAW

We discuss here only a few applications of Pascal law.

(a) Hydraulic Lift:

In a hydraulic lift or jack the pressure due to a small force F_1, say, applied at a piston of area A_1 ($A_2 > A_1$). Let F_2 be the force acting on the larger piston of area A_2. According to Pascal law the pressure p exerted on the piston of area A_1 is transmitted undiminished to the larger piston through the incompressible fluid in the channel. Thus

$$p = \frac{F_1}{A_1} = \frac{F_2}{A_2}$$

or $$\frac{F_2}{F_1} = \frac{A_2}{A_1}$$

$$F_2 = \frac{A_2}{A_1} \times F_1 \qquad (10.8)$$

As clear from the above equation

$F_2 > F_1$ since $A_2 > A_1$

Thus a small force acting on a small area A_1 generates a large force F_2 acting on a large area A_2. Under the action of this large force, the larger piston begins to move upward. Hence a car placed on the moving piston which acts as a platform is lifted as shown in Fig (10.7).

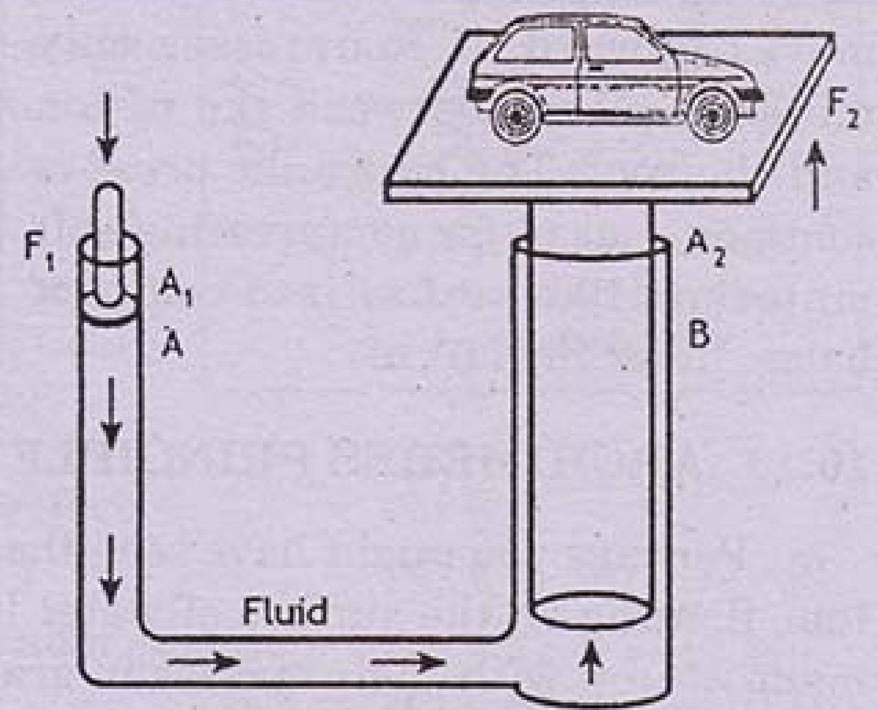


Fig. 10.7: A hydraulic lift is based on the fact that the pressure due to a small force applied at a narrow tube is equal to the pressure caused by a large force applied at a wide tube.

b. Hydraulic Brakes:

The working of hydraulic brakes is based on the principle of Pascal law. These brakes are used in automobiles. It consists of a tube which contains oil called brake oil. One end of the tube is fitted with a piston working in a master cylinder. There are in all four tubes through which the master cylinder is joined to four small cylinders (each containing a piston of large area) one for each wheel of the automobile. These small cylinders are called brake cylinders. When the break pedal is pressed by applying a force, the piston in the master cylinder begins to move inside the tube and exerts pressure on the oil. This pressure is transmitted to the oil in the brake cylinder. The undiminished transmitted pressure pushes the piston in brake cylinder which acts as a brake shoe attached to a calliper. The resulting friction stops the wheel. When the pressure on the pedal is released, the spring which connects the two brake-shoes contracts and pulls them off from the rotor. Thus the wheel is again free to rotate . Refer Fig (10.8)

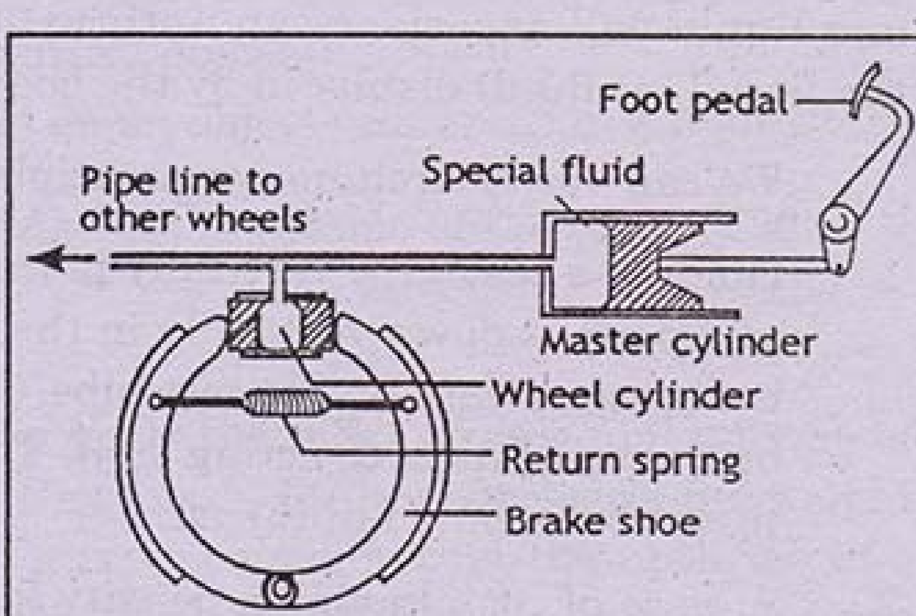


Fig. 10.8: Essential parts of a hydraulic brakes and their working

c. **Hydraulic Press:**

The working of hydraulic press is similar to that of a hydraulic lift. In a hydraulic press the piston of large cross sectional area used to lift an object is provided with a rigid roof over it. When the piston B of the large area moves upward, it compresses only material placed between the piston and the roof. The hydraulic press is commonly used for compressing soft materials like cotton into compact bales. Refer Fig (10.14)

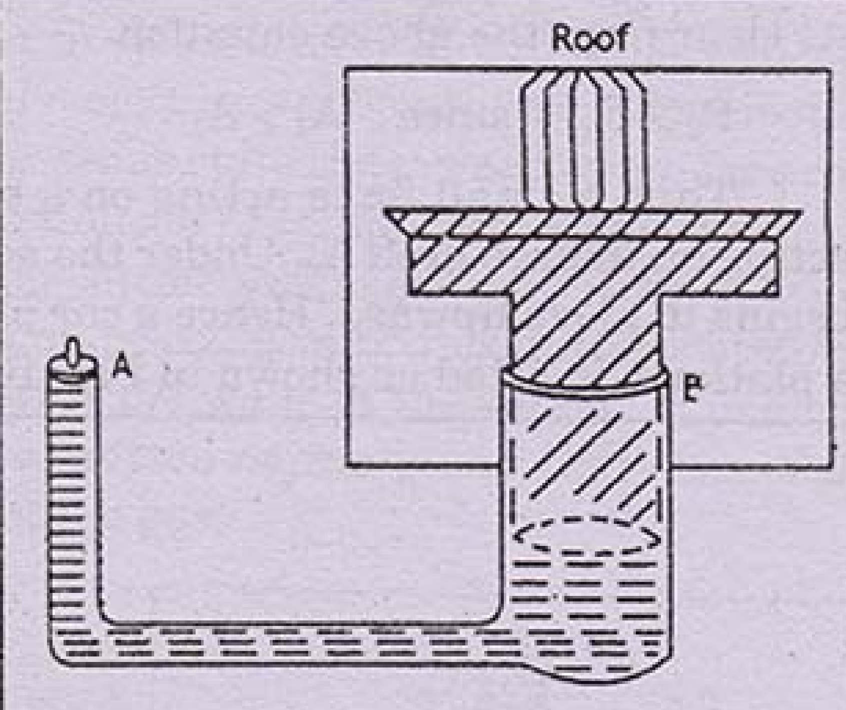


Fig. 10.9: Hydraulic Press

10.10 ARCHIMEDES PRINCIPLE

Perhaps you might have seen the ships made of iron with their masses in tons floating on the surface of water in oceans and seas and the needles also made of iron with their masses in grams and milligrams sinking in water. It is very surprising that a heavy body floats and a light body sinks. This means that for floating or sinking of a body it is not its mass which is important. It is something else which is important. The secret of the so called unlogical behaviour of floating and sinking bodies was made out in 300 B.C by a Greek scientist Archimedes. He discovered a principle regarding the floating or sinking of a body. Here we give the statement of Archimede's principle.

Statement:

The principle states that when a body immersed completely or partially in a liquid will experience an upthrust (upward force) equal to the weight of the liquid (or fluid) displaced by the body.

We can verify Archimedes principle from following examples:

1. Holding a test tube vertically place its closed end on the surface of water and push it down vertically in the water. You would feel as if some force is acting upward on the test tube. The deeper you push it, the greater will be the upward force acting on it. This upward force of water acting on the test tube is called upthrust.

2. A piece of cork having a density less than that of water needs to be held in water, if it is to remain in it. When released the cork will rise to the surface and will float. It is the upthrust (upward force) causing it to rise.

10.11 ANALYTICAL TREATMENT OF ARCHIMEDES PRINCIPLE

A liquid of density ρ is taken in a vessel. Consider a liquid of density ρ in a vessel. A cylindrical object of height 'h' and area of cross section 'A' is immersed in the liquid as shown in Fig. (10.10). The pressure exerted on the top surface of the object is ρgh_1 and that on the bottom surface is ρgh_2, where h_1 and h_2 are the depths of the top and bottom surfaces respectively. Let F_1 and F_2 be the forces due to pressure acting on the top and bottom surfaces respectively. As it is obvious that the force F_1 acts vertically downward where as the force F_2 acts vertically upward.

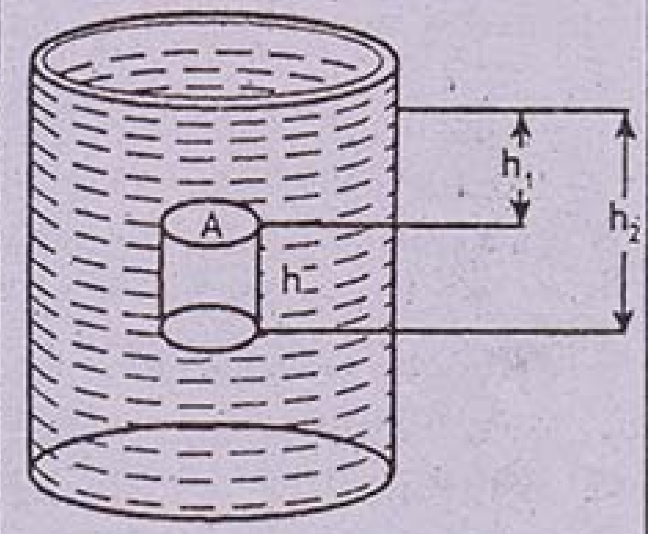


Fig. 10.10: A cylindrical vessel ful of liquid with a cylindrical object

The resultant vertical force on the top surface $= Ap + F_1$

$$= Ap + A\rho gh_1 \quad (10.9)$$

Resultant vertical force acting on the bottom surface $= Ap + F_2$

$$= Ap + A\rho gh_2 \quad (10.10)$$

Where p is the atmospheric pressure.

The net force acting on the object in the upward direction

$$= Ap + A\rho gh_2 - (Ap + A\rho gh_1)$$

$$= Ap + A\rho gh_2 - Ap - A\rho gh_1$$

$$= A\rho g(h_2 - h_1)$$

$$= A\rho gh$$

$$= AhPg \quad (10.11)$$

Where $h_2 - h_1 = h$ = height of the cylindrical object.

Thus "$Ah\rho g$" is the upthrust of a liquid on an object immersed in a liquid.

In the eq.(10.11) the expression "Ah" is the volume of the object. Call it V so that

$$Ah = V$$

Thus in the same equation $Ah\rho = V\rho$ is the mass of an equal volume of liquid displaced by the object. Hence the expression "$Ah\rho g$" is clearly the weight of the liquid displaced by the object. Thus we conclude that the upthrust of a liquid on an object immersed in it is equal to the weight of the liquid displaced by the object

Upthrust = Weight of the liquid displaced by the body.

Since an object immersed in a liquid experiences an upthrust, it has an apparent weight in the liquid given by

Apparent weight = Actual weight – weight of the displaced liquid

This means an object loses its weight in a liquid. This is the reason why it is easier to lift an object while in water than when it is above water.

10.12 BUOYANCY AND LAW OF FLOATATION

Buoyancy is the property of a liquid (or fluid) by virtue of which it exerts an upward force (or upthrust) on a body which is wholly or partially immersed in the liquid. This upward force is called buoyant force. In fact this buoyant force is equal to the weight of the liquid displaced by the body.

When an object is lowered in a liquid, it starts sinking down, displacing more and more liquid. As a result the buoyant force also begins to increase till it attains maximum value when the object is completely immersed into the liquid. When an object is immersed partially or completely in a liquid, it is acted upon by two forces (i) the weight of the object and (ii) the buoyant force acting vertically upward. The weight of the object is fixed but buoyant force is not fixed. The maximum value of the buoyant force occurs when the object is completely immersed. Its value depends on the volume (or mass) of the liquid displaced, that is, in every case its value is equal to the weight of the liquid displaced and it is further equal to the weight of the object.

If the maximum value of the buoyant force (F_{buoy}) max is equal to the weight of the body, that is,

$(F_{buoy})\ max = W,$

then the body will float in the liquid in such a way that upper surface of the object will coincide with the surface of the liquid. If the object is partially immersed so that the existing buoyant force acting on it is one half of its maximum value that is

(i) $F_{buoy} = (F_{buoy})\ max/2$

then,

$$F_{buoy} = \frac{(F_{buoy})\ max}{2} = W \text{ (weight of the object)}$$

If this is the case, then the object will float with half of its volume under the liquid

If the body is so immersed that

(ii) $\dfrac{(F_{buoy})\ max}{10} = W$

Then the object will float with one tenth of its volume within the liquid.

However if the body is so immersed that

(iii) $\frac{(F_{buoy})\max}{100} = W$

Then the body will float having 1% of its volume in the liquid. This means that if the existing buoyant force is small (as compared to the maximum buoyant force) then the portion of the object in the liquid will also be small.

If the existing buoyant force is greater than the weight of the object, that is:

$$F_{buoy} > W$$

then the object will float on the surface of the liquid with no portion inside the liquid. If such an object is forced to sink in a liquid by adding some loads on it, then it will rise again to the surface of the liquid after the removal of the loads. A cork floats on the surface of water because the upthrust of water on it is greater than its weight. A ship is so designed that its volume (space occupied by it) is very large. Hence the buoyant force which is equal to the weight of equal volume of water displaced by the ship is also very large and is greater than its weight. Hence a ship floats on the surface of water in a sea. A needle made of iron sinks in water because the buoyant force acting on it is less than its weight. It is due to the fact that the shape and the size of a common needle is so made that its volume is very small and hence the buoyant force which is equal to the weight of equal volume of water displaced by the needle is also small and less than the weight of the needle.

Submarines can float on the surface of water and when needed they can dive into water. They are fitted with large hollow ballast tanks. In order to dive in water they fill the tanks with water. This increases the weight of the submarine and hence submerges in water. The principle of the submarine is clear from the Fig. (10.11).

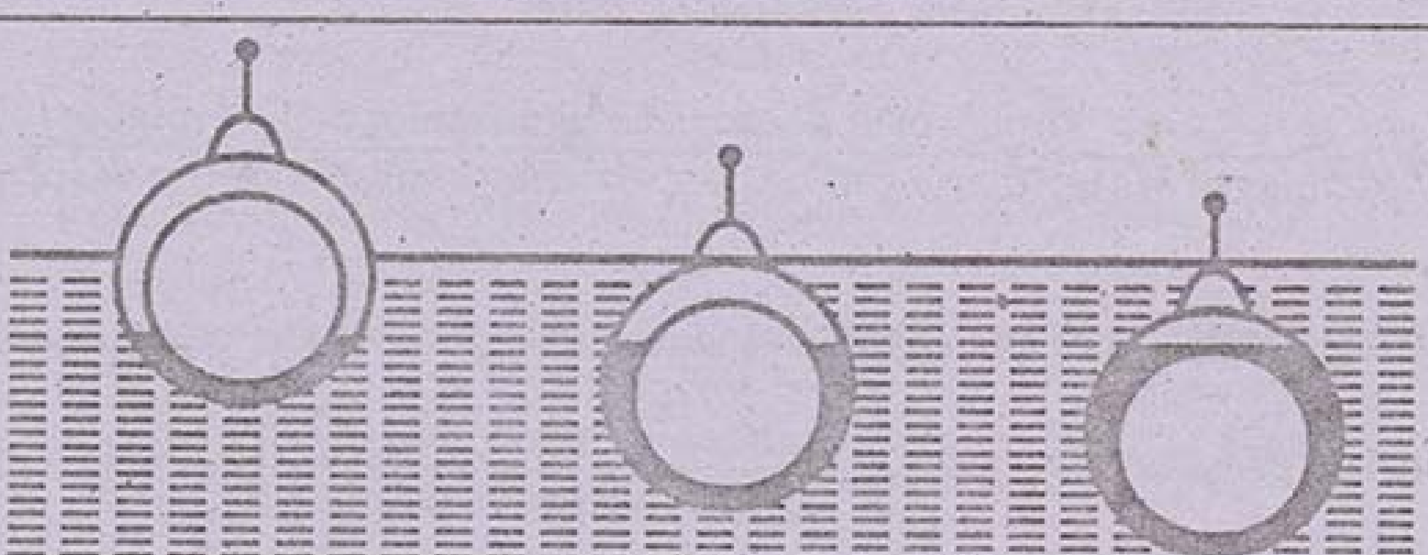


Fig. 10.11: Diving Submarine

To bring the submarine on to the surface extra load due to the water in the tanks is removed. This is done by forcing compressed air into the tanks which expels water from the tank.

A story about Archimede's finding:

According to a story, Hiero, the king of Syracuse, ordered a new crown. On receiving the crown, he was not satisfied. He doubted the gold smith of adulterating the gold in the crown with silver. The king asked his mathematician friend, Archimedes, if it were possible to determine if the crown were pure gold without cutting into it. Archimedes was contemplating the problem as he went to his bath. As he stepped into the full tub, he observed that the water which he displaced was equal to the volume of his body under the water. Archimedes shouted, "Eureka, Eureka"! ("I found, I found"), jumped out of the bath and ran down the street. What he had discovered was that the volume of the irregularly shaped crown could be found by immersing it in water. By comparing the weight of the crown with the weight of an equal volume of pure gold, he could determine if the crown were pure gold.

10.13 APPLICATIONS OF ARCHIMEDES PRINCIPLE

We describe here a few of applications of Archimedes principle.

(a) To find the specific gravity (relative density) of an insoluble solid substance (for example a piece of copper of suitable size):

Specific gravity (or relative density) of copper

$$= \frac{\text{density of copper piece}}{\text{density of water}}$$

$$= \frac{\text{mass of copper piece /volume of copper piece}}{\text{mass of water / volume of water}}$$

$$= \frac{\text{mass of copper piece /volume of copper piece}}{\text{mass of water displaced by copper piece / volume of the displaced water}}$$

Since

the volume of the displaced water = volume of the copper piece

$$\therefore \text{ specific gravity of copper } = \frac{\text{mass of copper piece (in air)}}{\text{mass of water displaced}}$$

$$= \frac{\text{weight of copper piece (in air)}}{\text{weight of water displaced}}$$

But weight of water displaced is equal to the loss of weight of copper piece in water, so

$$\text{specific gravity of copper} = \frac{\text{weight of copper piece (in air)}}{\text{loss of weight of copper piece in water}}$$

$$\text{specific gravity} = \frac{W_1}{W_1 - W_2} \quad \text{———} \quad (10.12)$$

where

W_1 = Weight of copper piece in air

W_2 = Weight of copper piece in water

From the above equation we can find the density of copper.

Since

$$\text{specific gravity of copper} = \frac{\text{density of copper}}{\text{density of water}}$$

$$\therefore \frac{\text{density of copper}}{\text{density of water}} = \frac{W_1}{W_1 - W_2}$$

$$\text{density of copper} = \frac{W_1}{W_1 - W_2} \times \text{density of water}$$

If d is the density of water at 4°C, then

$$\text{density of copper} = \frac{W_1}{W_1 - W_2} \times d \quad \text{———} \quad (10.13)$$

(b) Density of oil: (for example kerosene oil).

We have seen that the density of copper, using water, is given as

$$\text{density of copper} = \frac{W_1}{W_1 - W_2} \times \text{density of water} \quad \text{———} \quad (10.13)$$

Similarly the density of copper, using kerosene oil, is given as

$$\text{density of copper} = \frac{W_1}{W_1 - W_3} \times \text{density of kerosene oil} \quad \text{———} \quad (10.14)$$

Comparing equations (10.13) & (10.14), we have

$$\frac{W_1}{W_1 - W_3} \times \text{density of kerosene oil} = \frac{W_1}{W_1 - W_2} \times \text{density of water}$$

$$\therefore \text{ density of kerosene oil} = \frac{W_1/W_1 - W_2}{W_1/W_1 - W_3} \times \text{density of water}$$

$$\text{density of kerosene oil} = \frac{W_1 - W_3}{W_1 - W_2} \times \text{density of water}$$

$$\text{density of kerosene oil} = \frac{W_1 - W_3}{W_1 - W_2} \times d \quad \text{——— (10.15)}$$

Where W_3 is the weight of copper piece in kerosene oil.

(c) Density of cork:

Let W_1 be the weight of the cork in air.

Since cork is lighter than water, so its weight in water can not be determined directly. To find its weight in water a sinker is used. Let the weight of cork in air be W_1 and that of sinker in water be W_2. Also suppose that the weight of both of the sinker and cork in water be W_3. Hence the loss of weight of cork in water is $W_2–W_3$.

$$\text{density of cork} = \frac{\text{weight of cork in air}}{\text{loss of weight of cork in water}} \times \text{density of water}$$

$$\text{density of cork} = \frac{W_1}{W_2 - W_3} \times d \quad \text{——— (10.16)}$$

Where d is the density of water.

Example 10.5

An iceberg having a volume of 2060 cm^3, floats in sea water (density of seawater = 1.03g/cm^3) with a portion of 224 cm^3 above the surface of water.

Calculate the density of ice

Solution:

$V_1 = 2060cm^3$ = volume of iceberg

$V_2 = 2060cm^3 - 224cm^3 = 1836$ = volume of ice berg within water

$P = 1.03 \, g/Cm^3$ = density of sea water

$d = ?$ = density of ice

$W_1 = V_1 \times d \times g = 2060 \times d \times g$ = weight of iceberg

$W_2 = V_2 \times \rho \times g = 1836 \times 1.03 \times g$ = weight of displaced water.

Here W_2 is the upthrust of sea water on the iceberg.

According to Archimedes principle

$$\text{Upthrust} = \text{Weight of the ice berg}$$

$$W_2 = W_1$$

$$1836 \times 1.03 \times g = 2060 \times d \times g$$

$$\therefore d = \frac{1836 \times 1.03}{2060} = 0.918 \, g/cm^3$$

Example 10.6

A block of wood of volume $6.4 \times 10^{-6}m^3$ and of density $800Kg/m^3$ is floating in water. How much of wood is immersed in water? Density of water is $1000Kg/m^3$.

Solution:

$$\text{Weight of wood} = \rho Vg = 800 \times 6.4 \times 10^{-6} \times g$$

$$\text{Weight of displaced water} = \rho_1 \times V_1 \times g$$

$$= 1000 \times V_1 \times g$$

According to law of floating

$$\text{Weight of displaced water} = \text{weight of wood}$$

$$1000 \times V_1 \times g = 800 \times 6.4 \times 10^{-6} \times g$$

$$\therefore V_1 = \frac{800 \times 6.4 \times 10^{-6}}{1000} = 5.12 \times 10^{-6}m^3$$

10.14 KINETIC MOLECULAR THEORY OF MATTER

According to kinetic molecular theory, matter is composed of molecules. These molecules are always in motion which could be vibrational, rotational, translatory, or all of them hence the molecule could have all kind of K.E. Hence a molecule possesses kinetic energy due to vibration, translation or rotation. They are called vibrational, translational or rotational kinetic energies. The molecules of a substance attract each other with a force which depends on the distance between them. It decreases rapidly with the increase of distance between them.

The velocity of molecules depends on the temperature of the substance. The higher the temperature, the greater will be the velocity of molecules and lower the temperature, smaller would be the velocity of the molecules. Thus the kinetic energy of a molecule increases with the increase of temperature and decreases with the decrease of temperature of the substance.

10.15 SURFACE TENSION

Surface tension is the property of a liquid by virtue of which the free surface of a liquid behaves like a stretched-membrane tending to decrease the surface area.

Consider a free surface of a liquid. Draw an imaginary line AB on it. The molecules lying on one side of the line tend to pull away from the molecules lying on the other side as shown in Fig (10.12). Due to this action of the molecules the free surface of the liquid tends to decrease in area, producing a tension in it. Now we define surface tension as the force per unit length acting on either side of the imaginary line. The direction of this force is tangential to the surface and perpendicular to the imaginary line. If 'F' is the force acting on a length 'L', then the surface tension 'T' is given as

$$\sigma = \frac{F}{L} \qquad (10.17)$$

In the system international the unit of surface tension is N/m.

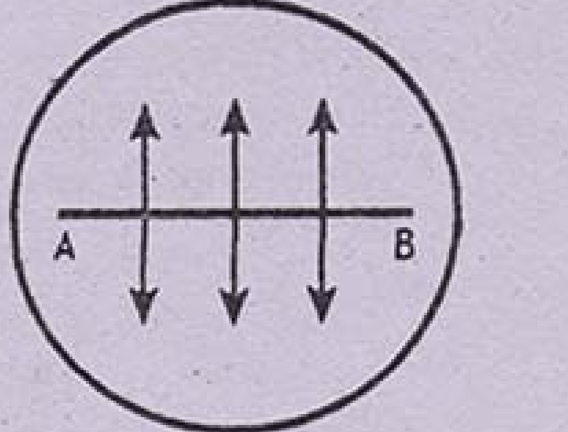


Fig. 10.12: Imaginary line drawn on the liquid. The molecules are pulling away from each other on the two sides of the line.

Explanation:

Consider a molecule 'A' of a liquid lying well inside it and a molecule 'B' on the surface of the liquid as shown in Fig (10.13). The molecule 'A' is attracted by all the molecules lying in its closest neighbour called the sphere of influence of radius 10^{-7}cm. The resultant force acting on this molecule is zero.

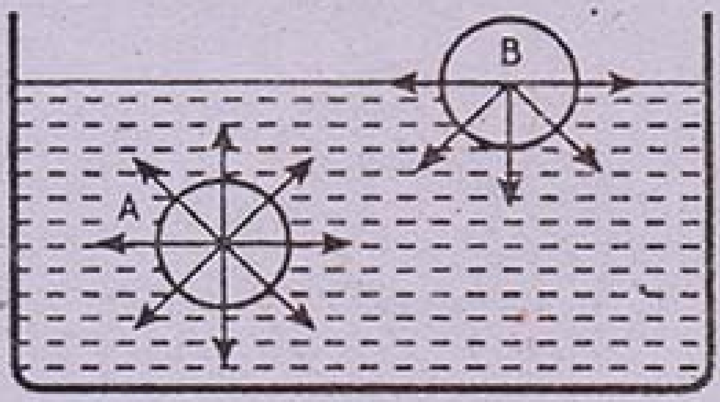


Fig. 10.13: Molecule A lying in the interior has no resultant force acting on it. The resultant force acting on molecule B lying on the surface is downward.

Now consider the molecule B. This molecule is acted upon by the molecules on the surface and those below the surface. Here the sphere of influence is half of the sphere as shown in Fig. (10.13). Due to the downward forces acting on the surface molecules, the free surface of the liquid behaves like a stretched membrane.

10.16 EXAMPLES SHOWING SURFACE TENSION

(a) In general a steel needle, if dropped in water, will sink because the density of steel is greater than that of water (Fig. 10.14a). However if the needle is made slightly oily and then placed on the surface of water in a horizontal position, it will float leaving a depression of water under it (Fig. 10.14b).

(a) (b)

Fig. 10.14

This means that the needle is not floating in ordinary sense. It is the surface tension of water which supports it and stops from sinking.

(b) A metallic ring is dipped into a soap solution and a light moist loop of cotton thread of any shape is gently placed over the soap film (Fig 10.15a). When the soap film within the loop is touched with a hot needle so as to break it, the loop takes a circular shape (Fig. 10.15b). Every point on the thread experiences a radially outward force of equal magnitude due to surface tension in the soap film out side the loop. These forces being due to surface tension lie in the plane of the soap film.

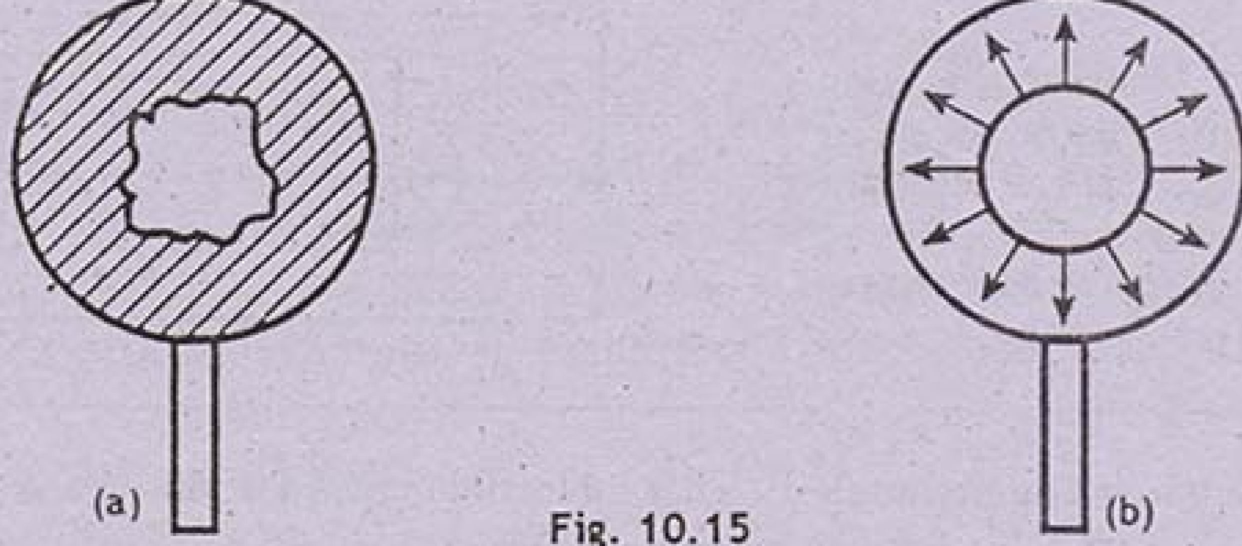


Fig. 10.15

(c) When a drop of olive oil is left gently inside a mixture of spirit and water (the density of the mixture being equal to that of olive oil) with the help of a pipette, the olive oil drop assumes a perfectly spherical shape (Fig. 10.16). Due to the surface tension, a liquid tends to keep its surface area minimum. For a given volume, the surface area of a sphere is minimum. Due to this reason the rain drops are spherical in shape.

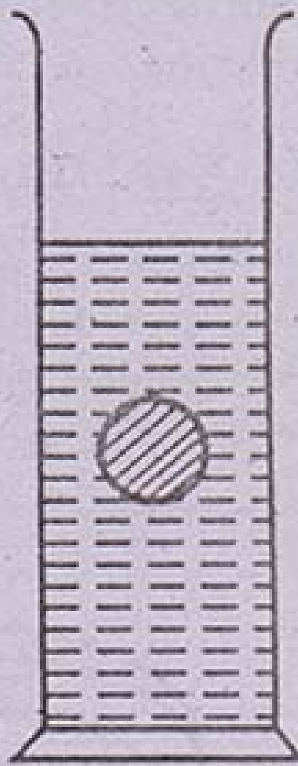


Fig. 10.16

(d) You may have observed water droplets falling from a tap, the drops are spherical in shape and is due to surface tension that minimizes the surface of the drop similarly. Molten lead when allowed to fall through the end of a narrow tube, the lead drops assume spherical shape due to surface tension. Lead shots are manufactured in this way in factories.

10.17 FLUID FRICTION (VISCOSITY)

In our daily life we observe that thin liquid, like water, alcohol, spirit etc. flow readily where as thick liquid, like coltar, castor oil, glycerine flow more slowly under similar conditions. The liquids of the second kind are said to be more viscous than the liquids of the first kind.

As there is friction between two solid surfaces tending to oppose their relative motion when one is made to slide over the other, there is also friction between two layers of even the same liquid (in general fluid) when they are in relative motion. This can be demonstrated by a simple experiment as described below.

Consider a liquid which flows on a smooth horizontal glass surface. We may suppose the liquid to be divided into different layers parallel to the fixed surface on which the liquid is flowing. It is found that different layers move with different velocities. For a streamline motion the layer in contact with the fixed surface is stationary. The velocity of the layers increases with the distance from the fixed surface in a perpendicular direction as shown in Fig (10.17). The length of the arrow gives the velocity of the corresponding layers.

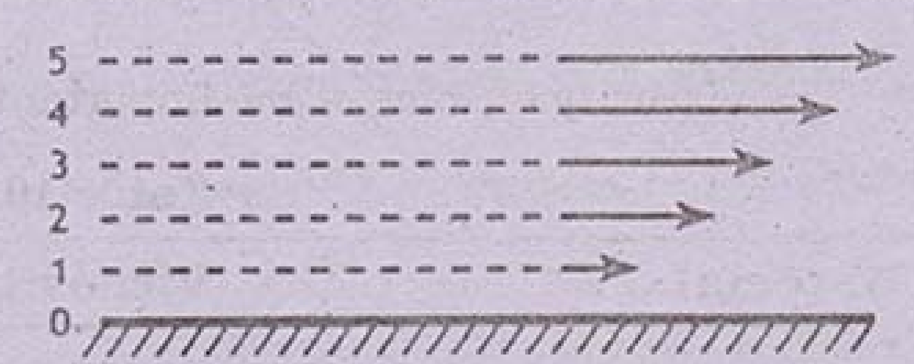


Fig. 10.17: Solid Surface

For any two layers, the upper layer moves faster than the lower one. Hence the upper layer tends to accelerate the lower layer which tends to retard the upper layer. Thus the two layers together tend to destroy their relative motion as if there is a backward dragging tangential force acting between the two layers. The property by virtue of which a fluid tends to oppose the relative motion between its different layers is called "viscosity" (or internal friction) of the liquid.

Consider any layer lying at a distance x from the stationary surface. Let 'A' be the area of the layer moving with a velocity v. It is found that the backward dragging force F acting on any layer is found to vary as:

$$F \propto A$$

$$F \propto v$$

$$F \propto \frac{1}{x}$$

$$\therefore \quad F = -\eta \frac{Av}{x} \qquad (10.18)$$

Where η is a constant depending on the nature of the liquid and is known as the coefficient of viscosity.

Negative sign shows that the dragging force acts in a direction opposite to the flow of the liquid.

From the above equation, the coefficient of viscosity is numerically equal to the retarding force required to maintain a velocity of 1m/s relative to the stationary layer. The unit of coefficient of viscosity is "poise" after Poisenille. The coefficient of viscosity of a liquid is said to be of 1 poise if the backward dragging force required to maintain a relative velocity of 1m/s between two layers each of area $1m^2$ and separated by a distance of 1m is 1N. Another commonly used unit of coefficient of viscosity is centipoise so that 1 centipoise = $\frac{1}{100}$ poise. We give here a table of viscosity of some important fluids.

The viscosity of some of the important fluids is given in Table 10.2.

Table 10.2

Material	**Viscosity (in centipoise) at 30°C**
Air	0.019
Acetone	0.295
Benzene	0.562
Ethanol	1.000
Methanol	0.510
Glycerine	629.000
Water	0.801

Viscous fluids have slow motion because viscosity which acts as a friction opposes the motion. The viscosity of a liquid is usually much small as compared to the friction between two solid surfaces. Hence an oil of high viscosity may be used as a lubricant. In heavy machines where there is considerable pressure on the bearings, a viscous lubricant should be preferably used because light or thin oil is easily squeezed out. For light machinery like sewing machines comparatively less viscous oil may be used as lubricant. Since the viscosity increases with a rapid rate with the decrease of temperature, lubricating may fail to form a protective layer at low temperatures. Hence while starting a vehicle during very cold weather it is advisable to allow the engine to run for some time so as to gain suitable temperature for driving.

SUMMARY

1. **Elasticity:** *The property of solids that restore them to their original shapes when external force ceases to act.*

2. **Stress:** *It is the force acting on unit area of an object.*

3. **Strain:** The change in the shape of an object caused due to stress.

4. **Hooke's Law:** Within the elastic limit of an object, the strain produced in an object is directly proportional to the applied force.

5. **Tensile Stress:** It is the stress that changes the length of an object.

6. **Young's Modulus:** Tensile stress divided by strain is called Young's Modulus.

7. **Pressure:** It is the force that acts normally on unit area of a surface.

8. **Pascal's Law:** If pressure is exerted on a liquid, the liquid distributes it equally in all directions.

9. **Archimedes Principle:** When an object is immersed in a liquid, the liquid pushes up that object with the force that is equal in magnitude to the weight of the displaced liquid.

10. **Buoyant Force:** It is the force that pushes an object immersed in a liquid in the upward direction.

11. **Kinetic Molecular Theory of Matter:** According to this theory the molecules of a gas are always in motion. The gas molecules attract one another with attractive force that decreases with the increase in the intermolecular distances.

12. **Surface Tension:** The intermolecular attractive forces give rise to a net force that acts along with surface of the liquid known as surface tension.

13. **Viscosity:** Viscosity is the force that arises due to the force of friction between different layers of a fluid in flow.

QUESTIONS

10.1 Answer the following questions.

(i) What do you understand by the term elasticity? Define stress, strain and yield point.

(ii) State and explain Hooke's law. Describe an experiment to verify Hooke's law.

(iii) What is meant by young's modulus of elasticity? How will you show that steel is more elastic than rubber?

(iv) What is atmospheric pressure? How will you measure it?

(v) State and explain Pascal law. Give some of its important applications.

(vi) What do you mean by viscosity? Define coefficient of viscosity. How does it depend on temperature?

(vii) Define surface tension. Explain it on the basis of kinetic theory of matter.

(viii) State and explain Archimedes principle. Give some of its applications in finding the densities of different substances.

10.2 Fill in the blanks.

(i) According to kinetic theory of matter, molecules are always in ________ motion.

(ii) ________ is the property by virtue of which a substance retains its shape and size after the removal of the stress.

(iii) (a) stress = ________ (b) longitudinal strain = ________

(iv) Elastic limit is defined as ________.

(v) In the field of elasticity 'Y' stands for ________.

(vi) Rubber is ________ elastic than steel.

(vii) The space above mercury in a simple mercury barometer contains ________.

(viii) The instrument used to measure atmospheric pressure is called ________.

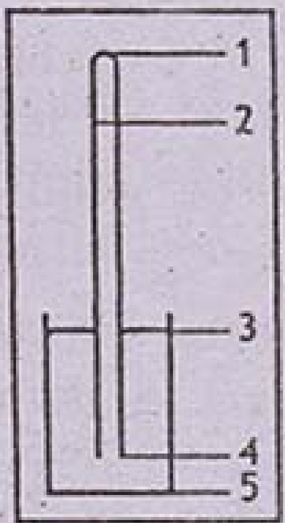
(ix) Surface tension is the property of a liquid by virtue of which its free surface appears like ________.

(x) The viscosity of a liquid increases with the ________ of temperature.

(xi) Viscosity increases with ________ of velocity.

(xii) According to Pascal law ________ is transmitted undiminished equally in all directions.

10.3 **Given below are a few possible answers to each statement of which one is correct. Identify the correct one.**



(i) The diagram shows a simple mercury barometer. The atmospheric pressure is given by the vertical distance between the following points.

a. 1 and 3 b. 1 and 4 c. 2 and 3 d. 2 and 5

(ii) A mercury barometer is set up in a plane. What happens to the mercury level if the plane gains altitude? a. It stays the same b. It rises higher c. It falls

(iii) A mercury barometer reads 76Cm. What will be the reading on a barometer filled with water at the same place? Mercury is 13.6 times denser than water.

(iv) Which one of the following does not cause the height of mercury column to vary?

a. Changes in atmospheric pressure.
b. Change in the value of g.
c. Change in the temperature of mercury.
d. Evaporation of mercury from the barometer reservoir.

(v) Elasticity of a substance depends on its:

a. temperature b. size c. nature

(vi) Archimedes principle is applied to determine:

a. specific heat b. specific gravity c. specific resistance

(vii) Rain drops are found spherical in shape because of one of the following properties of water.

a. surface tension b. viscosity c. pressure

(viii) Random motion of molecules in a fluid was first discovered by:

a. Robert Boyle b. Robert Brown c. Newton

(ix) An object appears lighter in water because of one of the properties of water.

a. buoyancy b. surface tension c. viscosity

(x) A liquid which is used for lubrication should be more:

a. transparent b. resistive c. viscous

(xi) The force of attraction between the molecules of the same liquid is:

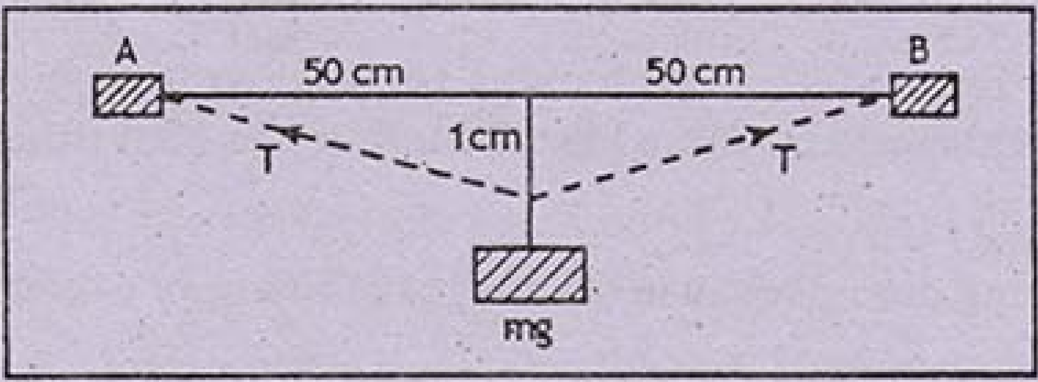
a. cohesive b. adhesive c. electromagnetic

PROBLEMS

10.1 A block of mass 0.1Kg is attached to a spring and placed on a horizontal frictionless table. The spring is stretched 20Cm when a force of 5N is applied. Calculate the spring constant.

(Ans. 25N/m)

10.2 A steel wire of diameter 0.8mm and length 1m is clamped firmly at two points A and B which are 1m apart and in the same horizontal plane. A block is hung from the middle point of the wire such that the middle point sags by 1Cm. Calculate the mass of the block.



Given $Y = 2 \times 10^{11} N/m^2$
$= 2 \times 10^{12}$ dynes/Cm^2

(Ans. 82 grams)

10.3 A steel wire 8m long and 4mm in diameter is fixed horizontally between two rigid supports. Calculate the increase in tension when the temperature falls by 10°C

$$\propto \text{ (for steel)} = 12 \times 10^{-6} K^{-1}$$
$$Y \text{ (for steel)} = 2 \times 10^{11} N/m^2$$

(Ans. 301.7 N)

10.4 A 2 Kg object is suspended from a copper wire 2.5m long and 0.6mm in diameter. How much will the wire stretch due to the gravitational force on the 2Kg object. Young's modulus for copper is $12.5 \times 10^{10} N/m^2$.

(Ans. 1.39×10^{-3}m)

10.5 A block of mass 92Kg and volume $0.031m^3$ lies at the bottom of the sea. How much force is needed to lift it? Take $1.03 \times 10^3 Kg/m^3$ as the density of sea water.

(Ans. 590N)

10.6 A canal lock gate is 20m wide and 10m deep. Calculate the thrust acting on it assuming that the water in the canal is in level with the top of the gate. Density of water is $1000Kg/m^3$.

(Ans. 9.8×10^6N)

10.7 A solid body floating on water has $\frac{1}{5}$ of its volume above the surface of water. What fraction of its volume will project above the surface, if it floats in a liquid of specific gravity 1.60. Take density of water = $1gm/cm^3$.

Ans. 0.5 $\left(\text{or } \frac{1}{5}\right)$

10.8 A cork of specific gravity 0.15 floats in water (sp. gravity 1.025) with 10CC above the surface. Calculate the mass of the cork.

(Ans. 1.757gm)

10.9 A body of density D is left free in a liquid of density d (D > d). Prove that the downward acceleration of the body while sinking in the liquid is given by the equation.

$$a = \left(1 - \frac{d}{D}\right) g$$

10.10 A block of wood (sp. gravity 0.85) floats on water. Some kerosene oil of sp. gravity 0.82 is poured on the surface of water until the wooden block is just immersed. Calculate the fraction of block lying below the surface of water in the second case.

$\left(\text{Ans. } \frac{1}{6}\right)$

CHAPTER – 11

HEAT

LEARNING OBJECTIVES:

- Temperature.
- Thermometric properties.
- General features of a thermometer.
- Liquid-in-glass thermometers.
- Mercury-in-glass thermometers.
- Thermal expansion.
- Expansion in solids.
- Expansion in liquids.
- Linear thermal expansion of solids.
- Volume thermal expansion.
- Applications of thermal expansion in solids.
- Bimetallic strip and its applications.
- Anomalous expansion of water.
- Effect of anomalous expansion of water.
- Expansion of gases.
- Boyle's law.
- Charle's law.
- General gas equation.
- Heat capacity.
- Specific heat capacity.
- Law of heat exchange.
- Measurement of specific heat capacity.
- Specific heat capacity of a liquid.
- Latent heat.
- Laws of fusion.
- Vaporization and latent heat of vaporization.
- Transmission of heat.
- Thermal conductivity.
- Practical application of conduction of heat.
- Practical applications of convection of heat.
- Applications of heat radiation.
- Thermos flask.

11.1 TEMPERATURE

The first physical quantity that we study in thermometry comes under the heading of heat. In thermometry we study about the measurement of temperature. Temperature is a thermal property of a body. This property is an easily measurable physical quantity.

Fluids always flow from a region of high pressure to that at low pressure. Electric charge (positive) flows from a body at higher potential to that at lower potential when electric contact is made between them. Electric current flows from higher potential to lower potential along a metallic wire. In a similar way heat flows from a body at higher temperature to that at lower temperature when they are brought in thermal contact.

Temperature describes the degree of hotness or coldness of a body. If we feel hot on touching a body, it is said to have degree of hotness, whereas if we

feel cold, we say that the body has a degree of coldness. The body having degree of hotness is said to be at a temperature higher than our body temperature. Similarly the body having degree of coldness is said to be at a temperature lower than our body temperature.

The touch method gives only a rough estimate of hotness or coldness. The exact degree of hotness or coldness is measured by a thermal device called thermometer.

11.2 THERMOMETRIC PROPERTIES

There are many physical quantities that change with the change of temperature. The change of a physical quantity with the change of temperature can be used to measure temperature. For this purpose the thermometer is accordingly calibrated. This property of a physical quantity is called thermometric property.

It is an experimental fact that liquid expand on increasing their temperature and contract on decreasing the temperature. Thus the change of volume of a liquid with temperature can be used to measure temperature.

Gases expand on increasing their temperatures and contract on decreasing the temperature. Hence the change in volume of a given amount of a gas at a constant pressure can be used to measure temperature. Similarly the pressure of a gas changes with change of temperature. Hence the change of pressure of a given mass of a gas at constant volume can be used to measure temperature.

Electrical resistances of metals change with temperature. Hence a change of resistance with temperature can be used to measure temperature.

11.3 GENERAL FEATURES OF A THERMOMETER

Every thermometer has two fixed points on it, the lower fixed point and the upper fixed point. These points are given arbitrarily assigned numerical values, which represent some fixed temperatures of water. The interval between these points is divided arbitrarily into divisions of equal width. Depending on the numerical values of these fixed points we have a thermometer of a particular type.

List of thermometers with their temperature dependent thermometric properties.

Table 11.1

S. No.	Thermometric property	Type of thermometer
1.	Volume variation	(i) Mercury-in-glass thermometer (ii) Alcohol-in-glass thermometer
2.	Volume variation at constant pressure	Constant pressure gas thermometer
3.	Pressure variation at constant volume	Constant volume gas thermometer
4.	Resistance variation	Platinum resistance thermometer

11.4 LIQUID-IN-GLASS THERMOMETERS

We discuss liquid-in-glass thermometers in some detail. Commonly there are two main types of liquid-in-glass thermometers. They are

(a) Mercury-in-glass thermometer
(b) Alcohol-in-glass thermometer

Here we give a list of main features of mercury and alcohol.

Mercury: It is a good conductor of heat. It expands uniformally on increasing its temperature and contracts on decreasing its temperature. Hence mercury is used in a liquid-in-glass thermometer

1. Freezing and boiling point:

Mercury freezes at –39°C and boils at 360°C.

Hence it can be used to measure a relatively long range of temperature.

2. Colour of mercury:

It has a shining silvery colour. Hence no colouring is needed to read the temperature in the thermometer.

3. Non-wetting property:

Mercury does not wet the glass surface of the thermometer tube. Hence the reading can easily taken. Since the meniscus is convex, the reading is taken at the top of the meniscus.

4. Thermal conductivity:

Since mercury has high thermal conductivity, it responds quickly to the change of temperature.

5. Thermal expansion:

Although mercury is a good conductor of heat, its thermal expansion is only one fourth of that of alcohol. Hence alcohol expands 4 times more than mercury for the same rise in temperature.

ALCOHOL

It has temperature dependent property. It expands on increasing the temperature by heating and contracts on decreasing the temperature by cooling. Hence alcohol is used as a thermometric substance. It has the following main features:

1. Freezing and boiling point:

Alcohol freezes at –115°C and boils at 78°C. Hence alcohol thermometers are not suitable for laboratory uses as the temperature encountered are usually alone the boiling point of alcohol. It can be used in polar regions where the temperatures are usually in the neighbourhood of –40°C.

2. Colour of alcohol:

It is a colourless liquid. It is usually coloured red when used in a thermometer.

3. Wetting property:

Alcohol wets the internal glass surface. When it contracts, it leaves same alcohol sticking to the capillary tube. Due to this property reading, cannot be taken clearly.

4. Thermal conductivity:

It has a low thermal conductivity. Hence it takes a long time as compared to mercury to attain the temperature of the surroundings.

5. Thermal expansion:

Its expansion is nearly four times the expansion of mercury for the same rise of temperature. Hence, the scale divisions can be made larger in size, making it easier to read.

11.5 MERCURY-IN-GLASS THERMOMETERS:

We limit ourselves to the study of mercury-in-glass thermometers.

Construction: We take a glass tube of a uniform and fine bore called the capillary of the thermometer. The capillary tubing ends on a glass bulb. The bulb and the capillary tube are filled with mercury. After heating to drive out all the air, the capillary tube is scaled. As the liquid (mercury) cools and contracts, the space above it in the capillary tube becomes practically a vacuum.

There are mainly three scales of temperature in which mercury is used as a thermometric substance. They are listed and described below:

(i) Celsius thermometer or Celsius scale of temperature:

This scale the temperatures are measured with respect to one standard temperature called the "Triple point" of water which is arbitrarily taken to be 273.16K (Note: Here K is pronounced as Kelvin). As an approximation, this temperature is taken as 273K. At this temperature all the three states of water (that is ice, water and water vapour) can coexist in equilibrium. This temperature is the zero of Celsius scale called 0°C. Here °C stands for degree centigrade. This is the lower fixed point of the thermometer. The upper fixed point on this scale is the temperature of steam at one atmospheric pressure which is taken to be 100°C. The interval between these points is divided into 100 equal parts. Each part measures 1°C.

Fig. 11.1: A simple mercury-in-glass thermometer

(ii) Kelvin scale of temperature:

The lower fixed point of this thermometer is the triple point of water which 273K. The upper fixed point is the temperature of steam at one atmospheric pressure. This temperature is marked as 373K. The interval between these points is divided into 100 equal parts. Each part measures 1K (one Kelvin, not one degree Kelvin). An interval of 1°C is equal to the interval of 1K. The zero of this scale marked 0K starts from –273°C. This is lowest temperature ever reached. This is the reason why this temperatures called absolute zero and the scale of temperature is also called absolute temperature scale or absolute scale of temperature.

From the above table it is clear that any temperature t°C measured in Celsius scale can be converted into Kelvin scale according to the relation

$$\text{or} \quad \left.\begin{aligned} t^\circ C &= (t + 273)K \\ TK &= T^\circ C + 273 \end{aligned}\right\} \qquad (11.1)$$

(iii) Clinical thermometer:

This thermometer is used by doctors to find the temperature of a patient. It is also found nearly in every house and is used to know the temperature of an ailing member of the family.

It is a modified form of an ordinary mercury thermometer. Since the normal temperature of a healthy person is about 37°C or 98.4°F, it has a short range of temperature from 35°C to 43°C or 95°F to 110°F.

This thermometer has a constriction in its capillary tube just above the bulb. On putting the bulb of the thermometer into the patients mouth, the temperature rises. The mercury in the bulb expands and passes through the constriction into the capillary tube. On cooling the mercury in the bulb contracts, but the mercury in the tube does not fall back to the bulb because the constriction prevents it from doing so. Thus the temperature of the patient as shown by the thermometer can be read at leisure.

In order to take fresh reading (the temperature of a patient) by the thermometer, the mercury thread is shaken back to the bulb.

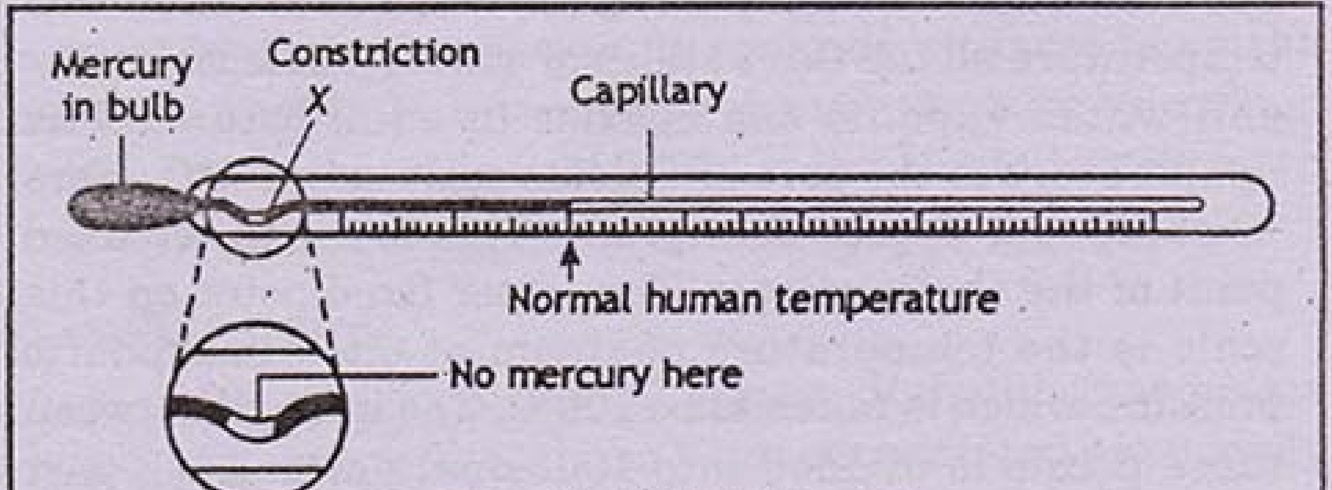


Fig. 11.1(a): Clinical Thermometer

(iv) Fahrenheit scale of temperature:

The lower fixed point on this scale is the triple point of water (or melting point of ice at one atmospheric pressure) which is marked as 32°F. the upper fixed point is the temperature of steam or boiling point of water at one atmospheric pressure. This temperature is marked as 212°F. the interval between these points is divided into 180 equal parts. Each part measures 1°F.

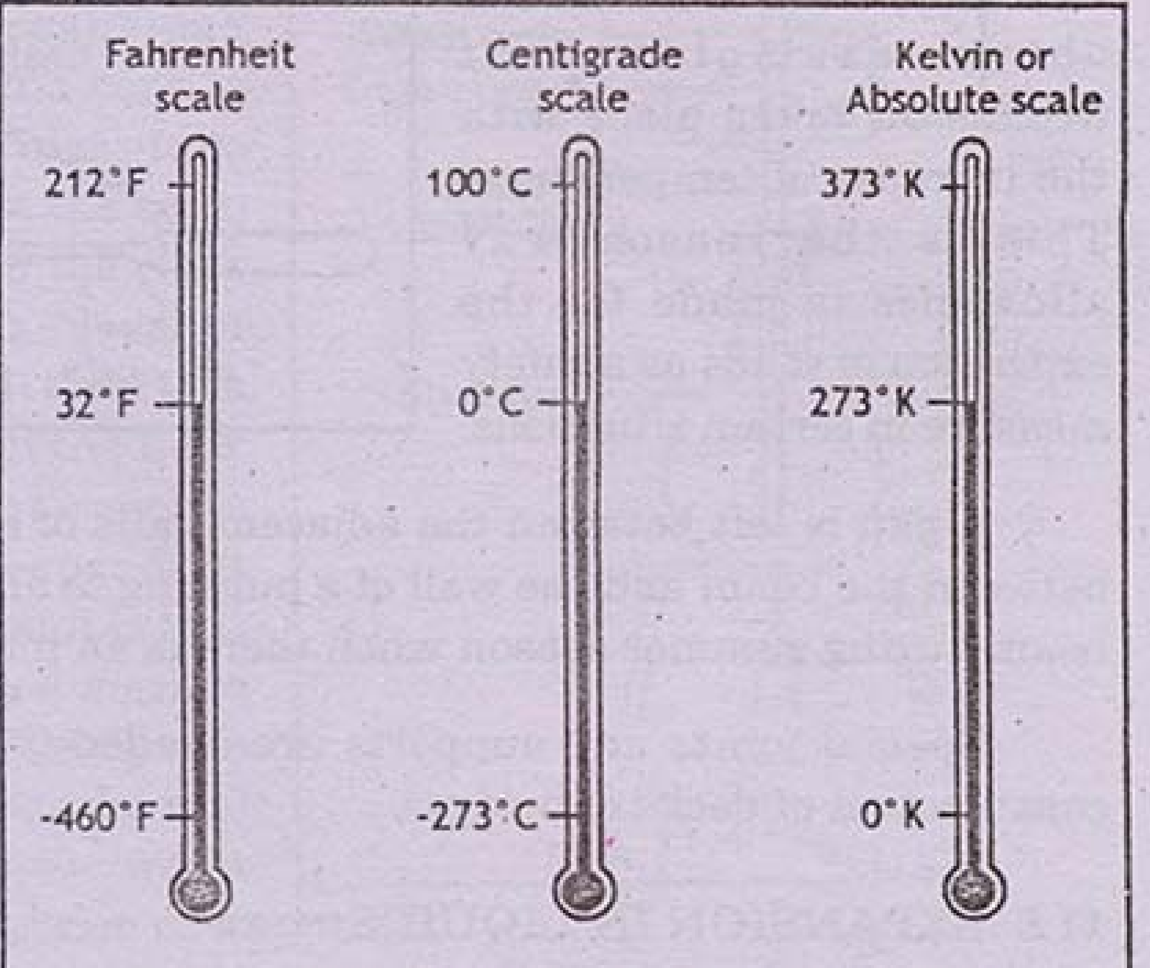


Fig. 11.1(b): three scales of temperature

11.6 THERMAL EXPANSION:

Almost all material bodies whether solid, liquid or gas expand on heating. This expansion is called thermal expansion.

11.7 EXPANSION IN SOLIDS:

A metallic lid tightly fixed on a jar is loosened by running hot water over it. As the temperature of the lid increases it expands more than the portion of the jar under the lid and hence loosening takes place.

If a hole is made in a metal plate and then heated the size of the hole increases because of its expansion. The inner edge of the plate forming the hole is metallic and hence expands on heating. This results in the increase of the size of the hole as clear from the Fig. (11. 2)

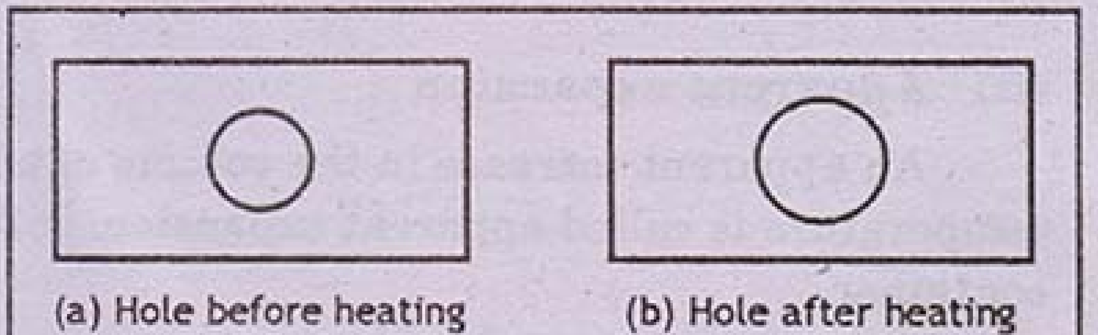


Fig. 11.2

Ring and ball experiment:

Take a metallic ball which just passes through a ring. The ball is heated through a relatively long range of temperature. The ball is again tried to pass but no use. This shows that the ball has expanded on heating.

It is clear from all the above examples that expansion takes place with the increase of temperature. This is the reason why allowance is made for the expansion of solids as a safety measure in certain situations.

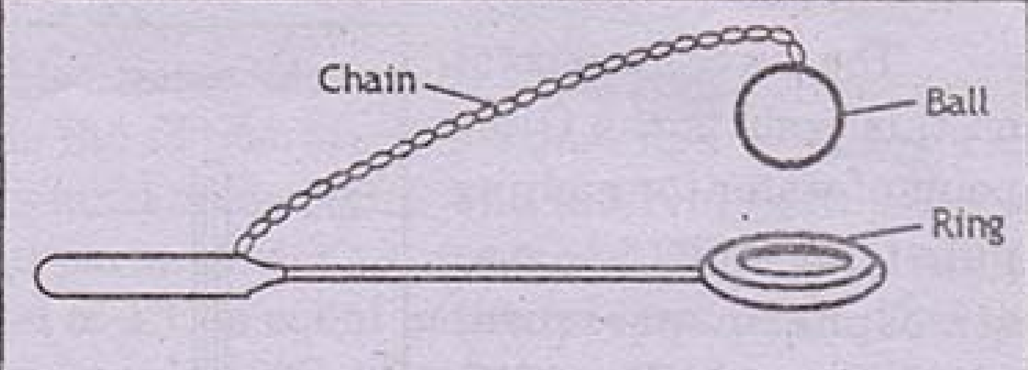


Fig. 11.3: Ball and ring apparatus

A gap is left between the adjacent rails of a railway track. Space is left between the beam and the wall of a building to allow the safe expansion of the beam during summer season when there is an increase in the temperature.

Special joints and supports are needed to allow the expansion in the construction of decks of bridges.

11.8 EXPANSION IN LIQUIDS:

A completely filled and tightly capped bottle of water cracks on heating. It is because of the more expansion of water than glass bottle on heating.

Temperatures are measured by liquid-in-glass thermometers using the property of thermal expansion of liquids.

There are two types of expansion associated with a liquid. They are (i) Real expansion and (ii) Apparent expansion.

(i) Real expansion

A real increase in the volume of a liquid, that takes place due to increase of temperature is called real expansion. This expansion is independent of the expansion of the container.

(ii) Apparent expansion

An apparent increase in the volume of a liquid that occurs due to rise of temperature is called apparent expansion. It depends on the expansion of the container.

When a liquid is taken in a container and heated, both the liquid and the container expand simultaneously. The difference of these expansions is called apparent expansion. If V_l and V_c are the expansion in volume of the liquid and the container respectively on heating, then the apparent expansion denoted by V_{app} is given as

$$V_{app} = V_l - V_c$$

or $$V_l = V_{app} + V_c$$

Here V_r is the real expansion of the liquid. We can demonstrate these expansions experimentally. A flask fitted with a cork is filled with coloured water. A narrow glass-tube is passed through the hole made in the cork. Water is filled to a mark 'a' made on the glass tube. The flask is then heated. The level of water in the glass tube first falls to a certain mark 'b' because of the expansion of the flask which is heated first. When heat reaches gradually to water, the latter also expands at a rate greater than the flask. The result is that the water level rises to a higher level marked as 'C', as shown in the Fig. (11.4). The volume of water in the tube from b to c gives the real expansion where as the volume of the water from a to c gives the apparent expansion of water.

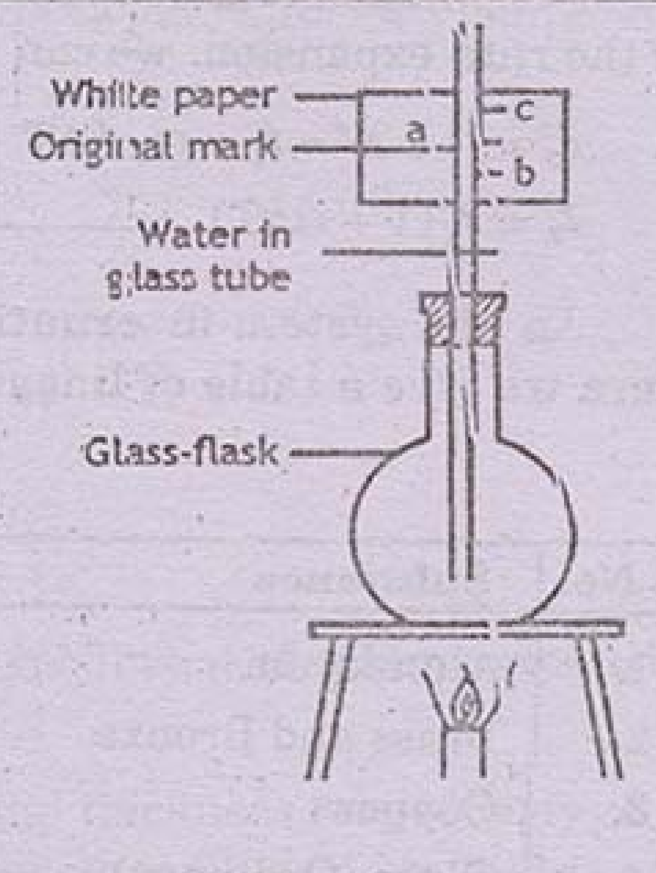


Fig. 11.4

11.9 LINEAR THERMAL EXPANSION OF SOLIDS:

One dimensional thermal expansion that is expansion along the length of a long solid body on heating is called linear thermal expansion.

Consider a thin rod so that its expansion is nearly one dimensional that is along its length. Let l_1 be its original length (before heating) at temperature T_1. The rod is heated to temperature T_2 so that its length increases to l_2 Δl be the increase in length ($\Delta l = l_2 - l_1$) when the rise in temperature due to heating is ΔT ($\Delta T = T_2 - T_1$). It is found experimentally that the increase in length varies directly with the original length and also varies directly with the increase of temperature that's

$$\Delta l \, \alpha \, l_1 \times \Delta T$$

$$\Delta l = \alpha \, l_1 \, \Delta T \quad \text{———————— (11.2)}$$

$$\alpha = \frac{\Delta l}{l_1 \, \alpha \, \Delta T} \quad \text{———————— (11.3)}$$

Where 'α' is a constant and depends on the nature of the material of the rod. It is known as the coefficient of linear expansion. It is expansion K^{-10} numerically increase in length per unit length per degree rise in temperature. This means that if we take a rod of length 1m and heat it through 1°C or 1K so that the increase in length is 24µm (2.4×10^{-6} m), then the coefficient of

linear expansion α is $24 \times 10^{-6}K^{-1}$. The greater the value of α, the greater will be the rate expansion. we can use eq.(11.2) to find ℓ_2.

$$\ell_2 = \ell_1 + \Delta\ell = \ell_1 + \alpha_1 \ell_1 \Delta T$$

$$\ell_2 = \ell_1 (1 + \alpha\Delta T) \quad \text{———————} \quad (11.4)$$

In the system international the unit of cost of linear expansion K^{-1}. Here we give a table of linear expansion of some common solids.

Table 11.2

S.No.	Substance		(K^{-1})
1.	Aluminium	24×10^{-6}	2.4×10^{-5}
2.	Brass and Bronze	19×10^{-6}	1.9×10^{-5}
3.	Copper	17×10^{-6}	1.7×10^{-5}
4.	Glass (Ordinary)	9×10^{-6}	0.9×10^{-5}
5.	Glass (Pyrex)	3.2×10^{-6}	0.32×10^{-5}
6.	Hard Rubber	80×10^{-6}	8.0×10^{-5}
7.	Ice	51×10^{-6}	5.1×10^{-5}
8.	Invar (Ni-Cr.alloy)	0.9×10^{-6}	0.09×10^{-5}
9.	Lead	29×10^{-6}	2.9×10^{-5}
10.	Steel	11×10^{-6}	1.1×10^{-5}
11.	Concrete	12×10^{-6}	1.2×10^{-5}

Example 11.1

A steel rod has a length of 10m at a temperature of 25°C. Calculate the increase in length if it is heated to 35°C. Given 'α' (for steel) = $1.1\times10^{-5}K^{-1}$.

Solution:

$\ell_1 = 10m$

$\Delta T = 35°C - 25°C = 10°C$

From eq.(11.2)

$\Delta\ell = \alpha \ell_1 \Delta T = (1.1 \times 10^{-5})(10m)$

$\Delta\ell = 1.1 \times 10^{-3}m = 0.0011m$

11.10 VOLUME THERMAL EXPANSION:

Three dimensional expansion that is a simultaneous expansion along three directions (along length, breadth and thickness or height) causing an expansion in volume on heating is called volume expansion.

Consider a block of volume V. Let the block be heated through a temperature ΔT so that the increase in volume is ΔV.

then $\ell t = \ell(1 + \alpha t)$

$$V = \ell_1^3 = \ell^3 (1 + \alpha t)^3$$

$$\Delta V \propto V \Delta T$$

$$\Delta V = ß\, V \Delta T \qquad (11.5)$$

$$ß = \frac{\Delta V}{\Delta T} \qquad (11.6)$$

Where ß is a constant and is known as a coefficient of volume expansion. Its unit is $°C^{-1}$ or K^{-1}.

Let ℓ_1, b_1 and t_1 be the length, breadth and thickness respectively of the block before heating let these quantities attain the values ℓ_2, b_2 and t_2 after heating through ΔT. We now calculate the increase in length, breadth and thickness of the block by considering its linear expansion along the three directions.

Let $\Delta \ell$, Δb and Δt be the respective increase on heating. From eq. (11.4)

$$\ell_2 = \ell_1 (1 + \alpha \Delta T)$$

$$b_2 = b_1 (1 + \alpha \Delta T)$$

and

$$t_2 = t_1(1 + \alpha \Delta T)$$

Where α is the coefficient of linear expansion.

From the above equations, we get

$$\ell_2 \times b_2 \times t_2 = \ell_1 \times b_1 \times t_1 (1+\alpha \Delta T)^3$$

$$V_2 = V_1(1+\alpha \Delta T)^3$$

$$V_2 = V_1(1+ 3\alpha \Delta T+3\alpha^2 \Delta T^2+\alpha^3 \Delta T^3)$$

Since α is small, higher ordered terms in α are neglected.

Thus

$$V_2 = V_1(1+ 3\alpha \Delta T)$$

or

$$V_2 = V_1+ 3\alpha \Delta T\, V_1$$

$$V_2 - V_1 = 3\alpha \Delta T\, V_1$$

$$\Delta V = 3\alpha \Delta T\, V_1 \qquad (11.7)$$

Comparing eqs.(11.5) and (11.7)

$$\beta \Delta T\, V_1 = 3\alpha \Delta T\, V_1$$

$$\beta = 3\alpha \Rightarrow \alpha = \frac{\beta}{3} \qquad (11.8)$$

Eq.(11.8) shows that the volume expansion is three times the coefficient of linear expansion.

11.11 APPLICATIONS OF THERMAL EXPANSION OF SOLIDS:

We have seen previously that the thermal expansion of liquids and gases is used to measure temperature by calibrating thermometers. In this section we will discuss briefly the applications of the thermal expansion of solids. In all these applications a bimetallic strip is used. A bimetallise strip is described below.

11.12 BIMETALLIC STRIP AND ITS APPLICATIONS:

Take two strips of dissimilar metals and weld them together. The strips in this form are said to form a bimetallic strip. When such a strip is heated, bending takes place because one strip expands more than the other. For example brass expands more than iron and so they can form a bimetallic strip as shown below:

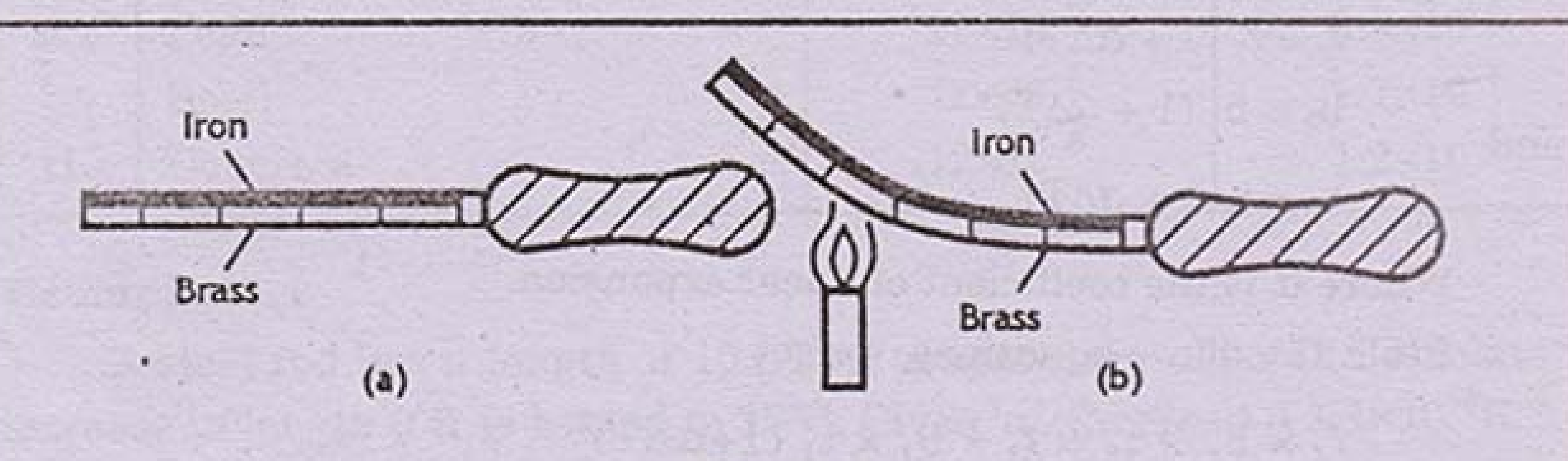


Fig. 11.5

There are many applications in which bimetallic strips are used. They are described briefly.

(i) Bimetallic thermometer:

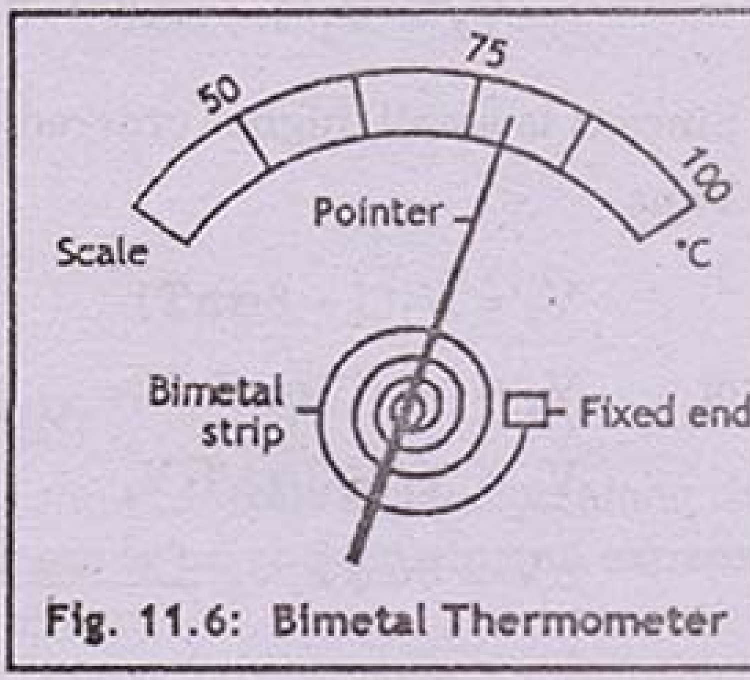


Fig. 11.6: Bimetal Thermometer

This thermometer consists of a bimetallic strip in the form of a long spiral. Its one end is fixed and the other and is firmly joined to a pointer which moves over a scale which is calibrated to measure temperature. When the temperature rises,

the spiral-turnings are more tightened because of the different amount of expansion of the strips forming the given bimetallic strip.

This results in the movement of the pointer on the calibrated scale which reads the temperature directly. The above details are shown in the Fig. (11.6).

(ii) Bimetallic thermostat:

A bimetallic strip can be incorporated to an electrical circuit to serve as a thermostat, a device to control the temperature. Suppose that a bimetallic strip is connected to an electric room heater. When the current flows through the heating element of the heater, its temperature rises and attains a value at which the bending of the bimetallic strip is so large that the electric contact is broken and the current ceases to flow. This results in the fall of temperature which reaches such a value that the bimetallic strip straightens to close the circuit again. The heating element is switched on and the bending of the bimetallic strip starts again the process of on and off is repeated and the temperature is controlled that is temperature range of the heater is fixed.

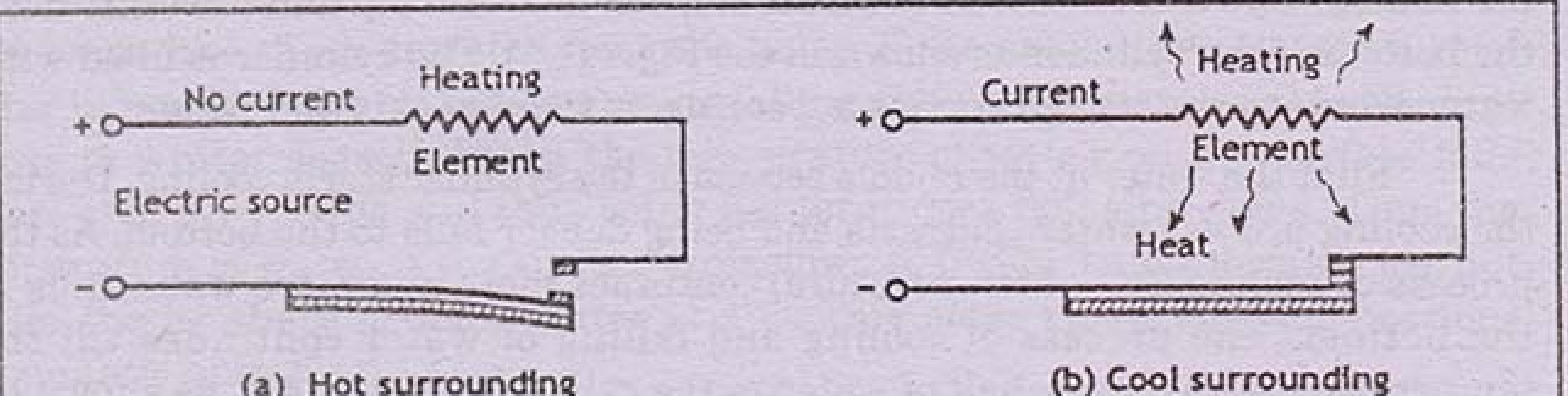


Fig. 11.7: Bimetallic Thermometer

(iii) Fire alarm:

A very useful application of a bimetallic strip is a fire alarm. One end of the bimetallic strip is fixed and the other is free. A 6 volt battery is connected between a metallic contact and the fixed end of the bimetallic strip through an electric bulb or an electric bell. The metallic contact is kept just above the free end of the bimetallic strip. All the details are shown in the Fig. (11.8)

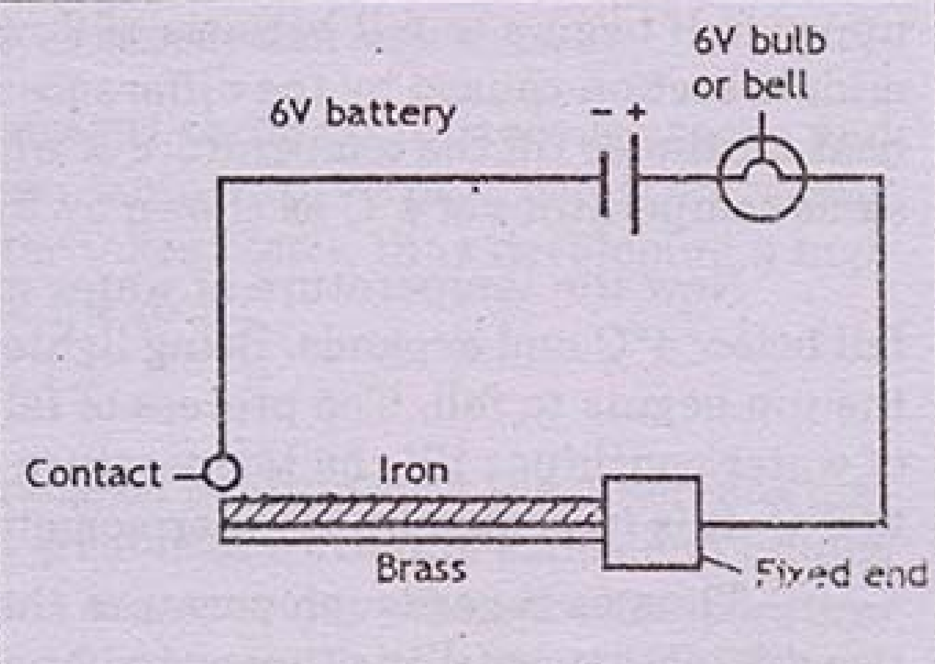


Fig. 11.8: Fire Alarm

When fire takes place, temperature rises. The bimetallic strip bends and touches the metallic contact. As a result current begins to flow in the circuit due to which the bulb glows or the bell begins to ring giving an alarming signal for fire.

11.13 ANOMALOUS EXPANSION OF WATER

When water is heated from 0°C to 4°C it contracts continuously instead of expanding. Conversely expands when cooled down from 4°C to 0°C. This expansion of water (from 4°C to 0°C) is called anomalous expansion.

When heated from 4°C to 100°C its expansion is normal like the expansion of other liquids. Since water contracts when heated from 0°C to 4°C and expands from 4°C to 100°C, its volume is smallest and the density is maximum at 4°C. Hence water is heaviest at 4°C.

The normal and anomalous behaviour of water with temperature is demonstrated by an experiment called Hopes' experiment as described and explained below.

A long metal cylinder is taken. It carries a circular trough around it in the middle. Two thermometers are inserted, one near the top and the other near the bottom of the cylinder as shown in the Fig. (11.9). The cylinder is filled with water whereas the trough carries a freezing mixture of salt and water.

First the water at the middle section of the cylinder starts cooling. During the cooling process water contracts and being denser falls to the bottom. As the process of cooling (fall of temperature) continues more and more water falls to the bottom. The process of cooling and falling of water continues till the temperature of the lower half of water in the cylinder reaches 4°C as shown by the thermometer T_2. Thus the temperature in the lower half is at 4°C and that in the upper half is above 4°C as shown by the thermometer T_1. Further fall of temperature and flowing down of water stops. Instead the temperature of the upper half begins to fall because of flow of heat from upper half towards the middle section caused by the difference of temperature. The process of flow of heat continues till the temperature of whole of water in the cylinder attains the same temperature of 4°C as shown by the thermometers T_1 & T_2.

Now the temperature of water at the middle of the cylinder begins to fall below 4°C and expands. Being lighter it goes up and so the temperature at the top begins to fall. The process of falling of the temperature and going up of water continues till the temperature of whole of the upper half cools down to 0°C. It is indicated by the thermometer T_1.

Thus as regards temperature the water in the cylinder is divided into two distinct parts (i) one (upper half) with temperature 0°C and (ii) the other (lower half) with temperature 4°C.

The normal and anomalous behaviour of water with temperature is also clear from the time-temp graph (Fig. 11.9).

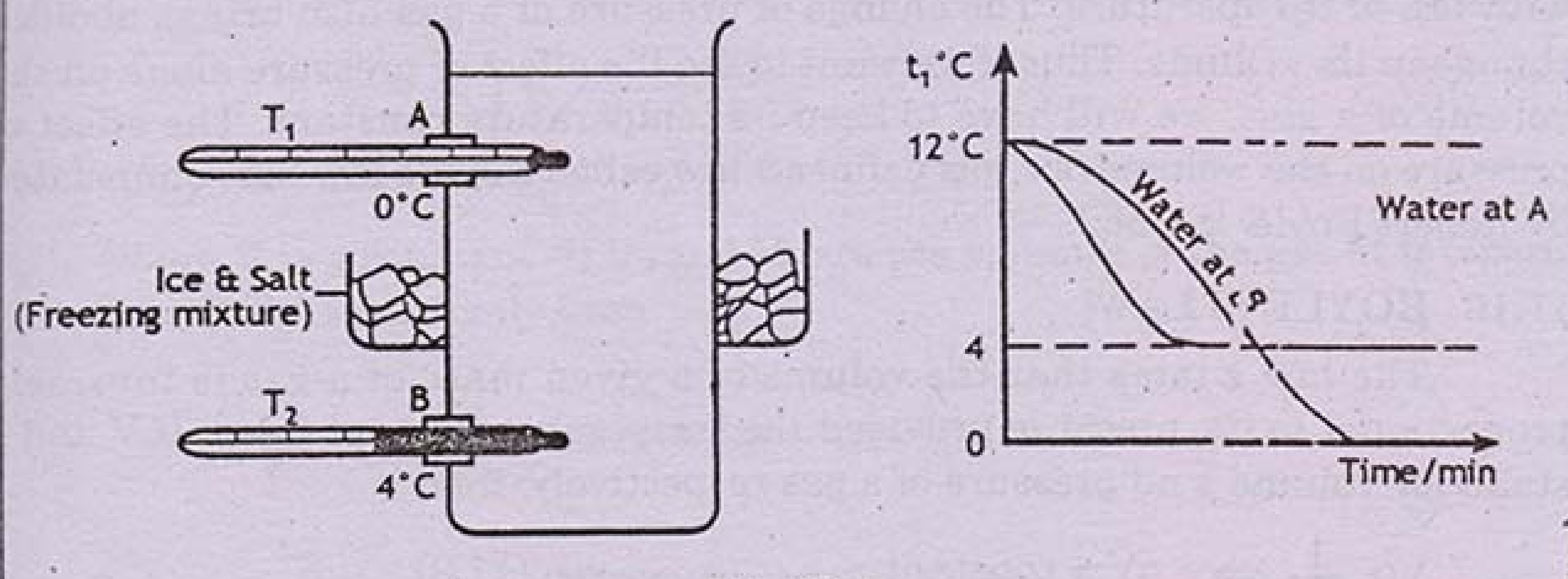


Fig. 11.9

11.14 EFFECT OF ANOMALOUS EXPANSION OF WATER

1.. Anomalous expansion of water is a blessing from God to marine life. In cold areas like those situated near the poles of the earth the temperature of water in ponds, lakes and oceans starts falling and goes down to 4°C in the first phase in winter season. When the temperature of water on the surface falls below 4°C, it expands and thus being lighter it does not sink down. Thus the temperature of water at the surface cools down to 0°C or below forming ice, but under the ice there is dense water at 4°C. This helps in serving the precious life aquatic animals.

2. In winter season the water supply pipes open to the atmosphere often burst when the temperature of the surroundings falls below 4°C. This is due to the fact that water below 4°C expands (anomalous expansion) and exirts pressure to the walls of the pipes and thus causes damage to it.

3. In rainy season a lot of water sweeps through numerous cracks and fissures in rocks. In winter season when temperature falls below 4°C going down to 0°C or below, expansion of water takes place, thus developing a high pressure. This results in the breaking of the rocks.

11.15 EXPANSION OF GASES:

Gases expands like solids and liquids on heating with the difference that the relative expansion of gases is much greater than that of solids and liquids for the same rise of temperature. On a hot summer day when the temperature is on the peak, the tube of a cycle bursts when placed in the direct sun. The air in the tube expands with the rise of temperature of the surroundings. Due to

expansion the air in the tube exerts pressure on the walls of the tube which bursts. This is an example of expansion of gases due to rise of temperature.

However gases expand not only with the rise of temperature and contract with fall of temperature. The change of pressure of a gas also brings about a change in its volume. Thus if we want to see the effect of pressure alone on the volume of a gas, we will have to keep its temperature constant. The effect of pressure on the volume of a gas defines a law called Boyle's law first enunciated by Robert Boyle in 1662.

11.16 BOYLE'S LAW:

The law states that the volume of a given mass of a gas is inversely proportional to its pressure provided the temperature is kept fixed. If V and p stand for volume and pressure of a gas respectively, then

$$V \alpha \frac{1}{p} \Rightarrow pV = \text{constant} \quad \text{-------------------} (11.9)$$

If a gas has different sets of pressure and its corresponding volume like (p_1, V_1), (p_2, V_2), ..., (p_n, V_n), then according to Boyles' law

$$p_1V_1 = p_2V_2 = \text{.........} = p_nV_n = \text{Constant} \quad \text{----} (11.10)$$

The value of constant in Boyles' law depends on the mass of the gas. If the mass of the gas also changes, it can be easily shown that

$$\frac{\Delta PV}{\Delta m} = \text{Constant}$$

Where m_1 and m_2 are the mass of the gas in the states 1 and 2 respectively.

A graph plotted between pressure and volume is shown in Fig. (11.10). It is a hyperbola. The graph shows that the volume varies with pressure in such a way that the product is constant.

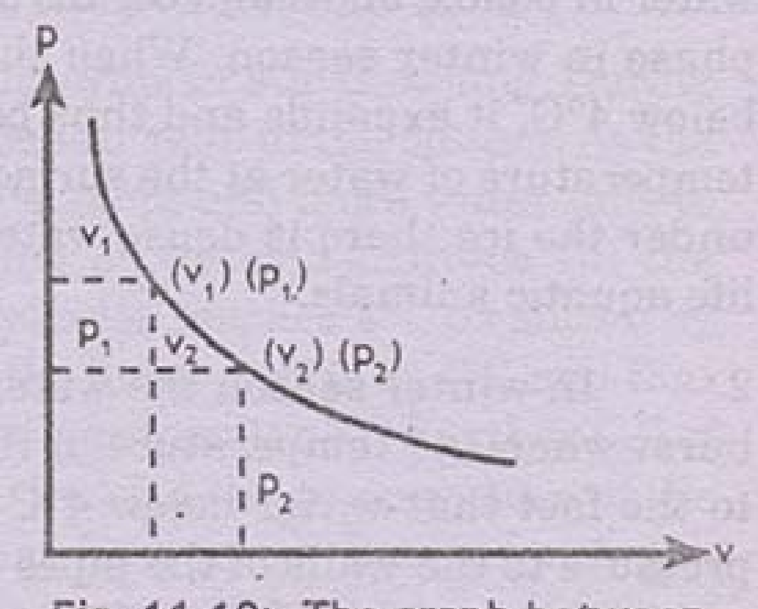


Fig. 11.10: The graph between pressure and volume is a hyperbola showing the inverse relation between them

11.17 CHARLE'S LAW:

Since the volume of a gas varies with its pressure (Boyle's law), so to see the effect of temperature on volume we keep the pressure constant. The exact relation between the temperature and volume of a gas is given by a law called charle's law first enunciated by Jacques Charles.

The law states that the volume of a given mass of a gas is directly proportional to its absolute temperature provided the pressure is constant. If V and T stand respectively for volume and temperature of the gas then

$$V \propto T$$

or $$V = KT$$

where K is a constant. If V_1 and V_2 are the volumes of the gas at temperatures T_1 and T_2 respectively then

$$\frac{V_1}{T_1} = \frac{V_2}{T_2} \quad (P = \text{constant}) \text{ -------- } (11.11)$$

Since the volume varies directly with temperature, the graph plotted between these quantities is a straight line as shown in Fig. (11.11). If the straight line obtained is extended, it cuts the temperature axis at –273°C, where the volume of the gas if existed is zero. A gas with zero volume is untengible concept. A gas with zero volume can not exist because gas is a kind of matter which occupies space. As a matter of fact the gas is liquefied much before reaching this point and hence Charle's law can not be applied. As we have discussed earlier –273°C which is equal to 0K on absolute scale of temperature is called absolute zero. Thus according to Charle's law absolute zero is the temperature at which the volume of a gas should be zero. Kinetic theory provides a better definition of absolute zero according to which this is the temperature at which all the molecules of a material body cease to move.

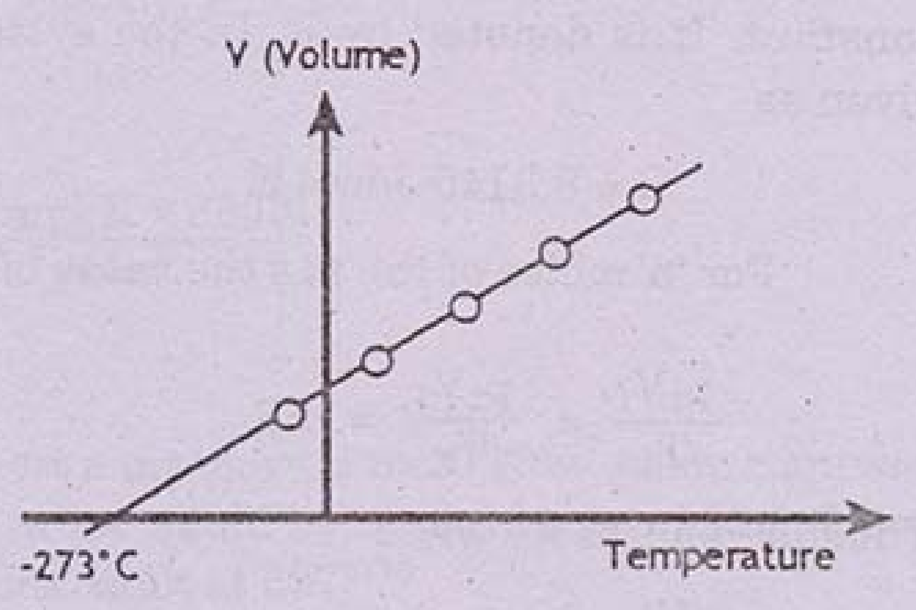


Fig. 11.11: Graph showing the direct relation between temperature (absolute) and volume

11.18 GENERAL GAS EQUATION:

So far we have discussed the expansion of gases under restricted conditions. One of the three state variables (temperature in Boyle's law and pressure in Charle's law) is kept fixed during expansion. We now change both

the temperature and pressure simultaneously and see their combined effect on volume. An equation relating all the three quantities (pressure, volume and temperature) is called the general gas equation. We will use boyle's law and Charle's law to obtain this equation. Let p, V and T be the pressure, volume and temperature of a gas. According to Boyle's law

$$V \propto \frac{1}{p} \quad \text{(when temperature is constant)}$$

According to Charle's law

$$V \propto T \quad \text{(when pressure is constant)}$$

$$\therefore V \propto \frac{T}{p} \Rightarrow \frac{pV}{T} = \text{constant}$$

If (p_1, V_1, T_1) and (p_2, V_2, T_2) define two states of a gas, then

$$\frac{p_1V_1}{T_1} = \frac{p_2V_2}{T_2} = \text{constant} \quad \text{--------------- (11.12)}$$

The value of constant depends on the mass of the gas expressed in moles. For one mole of the gas the constant has a particular name the universal gas constant. It is denoted by R. In the system international the value of R is given as

$$R = 8.3145 \text{ J/mol.K}$$

For 'n' moles of the gas the value of the constant is nR. Hence we have

$$\frac{p_1V_1}{T_1} = \frac{p_2V_2}{T_2} = nR$$

Thus in general we write

$$\frac{pV}{T} = nR \Rightarrow pV = nRT \quad \text{---------------- (11.13)}$$

This the familiar form of the general gas equation.

Example 11.2: An insulated cylinder fitted with a piston contains oxygen at a temperature of 20°C and a pressure of 15 atm in a volume of 22 litres. The piston is lowered down, decreasing the volume of the gas to 16 litres and the temperature is raised to 25°C. Assuming oxygen to be have like an ideal gas, what is the final pressure of the gas?

Solution:

From general gas equation.

$$\frac{p_1V_1}{T_1} = \frac{p_2V_2}{T_2}$$

Here

$p_1 = 15$ atm, $V_1 = 22$ litres, $T_1 = 20+273=293K$

$T_2 = 25+273 = 298K$, $V_2 = 16$ litres

$P_2 = ?$

$$P_2 = \frac{p_1V_1T_2}{V_2T_1} = \frac{15 \times 22 \times 298}{16 \times 293} = 21 \text{ atm}$$

Example 11.3: Calculate the volume occupied by 2 moles of a gas at 7°C and a pressure of 1 atm = $1.01 \times 10^5 N/m^2$.

Solution:

$p = 1.01 \times 10^5 N/m^2$ = pressure of the gas
$T = 7°C = (7 + 273)K = 280K$ = temperature
$n = 2mol$ = number of moles
$R = 8.31$ J/mgl.K = Universal gas constant
$V = p$ = Volume occupied by the gas
From general gas equation

$$pV = nRT$$

$$V = \frac{nRT}{p} = \frac{2\text{mol} \times 8.31 \text{ J/mgl.K} \times 280\text{K}}{1.01 \times 10^5 \text{ N/m}^2}$$

$$= 0.046 \text{ m}^3$$

Example 11.4: If 4 moles of a gas exert a pressure of $9\times10^4 N/m^2$ when confined in a tank of $40m^3$ capacity at 127°C, what would be the pressure of 200 mol of the same gas when confined in a $100m^3$ tank at 527°C?

Solution:

n_1 = 4mol = Initial amount of gas
p_1 = $9 \times 10^4 N/m^2$ = Initial pressure
V_1 = $40m^3$ = Initial volume
T_1 = (127+273)K = 400K = Initial temperature
p_2 = ? = Final pressure
V_2 = $100m^3$ = Final volume
T_2 = (527 + 273)K = 800K = Final temperature
n_2 = 200 mol = Final amount of the gas

Using general gas equation as

$$\frac{p_1V_1}{n_1T_1} = \frac{p_2V_2}{n_2T_2}$$

$$P_2 = \frac{p_1 \times V_1 \times T_2 \times n_2}{n_1 \times p_2 \times T_1 \times V_2}$$

$$= \frac{9 \times 10^4 Nm^{-2} \times 40m^3 \times 800K \times 200}{4 \times 400K \times 100m^3}$$

$$P_2 = 3.6 \times 10^4 N/m^2$$

11.19 HEAT CAPACITY:

Take a number of samples of different amounts of the same substance (say, copper) and heat them through the same range of temperature. It will be found that the heat required for this purpose is different for different samples. If ΔQ is the amount of heat needed to raise the temperature of a sample through ΔT, then the ratio $\frac{\Delta Q}{\Delta T}$ is known as the heat capacity of the sample (not of the substance). it is denoted by C so that

$$C = \frac{\Delta Q}{\Delta T} \quad \text{.................................} \quad (11.14)$$

In the system international the unit of heat capacity is JK^{-1}

Example 11.5: 0.5 kg of copper needs 1950J of heat to raise its temperature through 10°C. Calculate the heat capacity of the sample.

Solution:

ΔQ = 1950J = heat needed
ΔT = 10°C = rise in temperature
C = ? = heat capacity

Formula used

$$C = \frac{\Delta Q}{\Delta T} = \frac{1950J}{10°C} = 195JC^{-1}\ (195JK^{-1})$$

11.20 SPECIFIC HEAT CAPACITY:

Take equal amounts of different substances for example, copper, iron and water. Heat them for a given interval of time under the same flame so that they all absorb the same amount of heat energy from the source. At the end of the interval you will find that the rise in temperature is not the same for all.

It is maximum for copper, minimum for water and intermediate for iron (although much greater than that for water). This shows that copper is heated quickly where as water very slowly.

This means that if they are all heated so as to have the same rise in temperature, copper will need least amount of heat and the water the greatest one for this purpose. The amount of heat needed depends on the nature of the substance in addition to its logical dependence on the mass and the rise in temperature of the substance. It is found experimentally that the amount of heat required is directly proportional to both the mass and the rise in temperature of the substance. Thus if ΔQ is the amount of heat added to raise the temperature of m kg of a substance through ΔT, then

$$\Delta Q \propto \Delta T$$
$$\propto m$$
$$\therefore \quad \Delta Q = cm\Delta T \quad \text{----------------------} \quad (11.15)$$
$$\text{or} \quad c = \frac{\Delta Q}{m\Delta T} \quad \text{----------------------} \quad (11.15a)$$

Where c is a constant and known as the specific heat capacity of the substance. It depends on the nature of the substance and is entirely independent of its mass and the rise in temperature. If c is small for a substance, the heat needed will also be small and if it is large, the heat needed will also be large under the similar condition of mass and rise in temperature for all substance. In the system International the unit of specific heat capacity (or specific heat) is $Jkg^{-1}K^{-1}$.

From eq.(11.15)

$$mc = \frac{\Delta Q}{\Delta T}$$

or $$mc = C$$

where $C = \frac{\Delta Q}{\Delta T}$ is the heat capacity as defined earlier

Thus

$$C = mc \quad \text{------------------------------} \quad (11.16)$$

where c is specific heat.

The heat needed for this purpose will depend solely on their specific heats. The substance with large value of specific heat will need more heat as compared to that having low specific heat. For example if we take 1kg of water and copper each and heat them through 1K, then the heat energy needed for water is nearly 11 times $\left(\frac{4200}{390}\right)$ as large as for copper. In other words this

means that if we give same amount of heat energy to water and copper having same mass, then the rise in temperature of copper will be more than that of water. This further means that copper will be heated quickly as compared to water. Thus we conclude that under identical condition a substance with low specific heat will be heated quickly as compared to the substances having large values of specific heats.

We see that water has large heat capacity as compared to other substances. This means that water needs a lot of heat energy to warm up. A warm water has a good store of internal energy.

Similarly the loss of a large amount of heat energy causes only a small drop in the temperature of water. This is the reason why the temperature of the sea rises and falls very slowly. The daily temperature variations is thus smaller near coastal areas. For the same reason coastal areas and is lands keep cooler in summer and warmer in winter compared to inland areas.

Thus due to high specific heat capacity of water the coastal areas have moderate climate through out the year.

The high specific heat of water (as well as its cheapness and availability) accounts for its use as the circulating liquid in central heating systems and as a cooling liquid in car engines. Another common application is the use of hot water bottles to keep warm. This is due to ability of hot water to store a large amount of energy.

11.21 LAW OF HEAT EXCHANGE:

The law was first enunciated by Prevost. According to this law when two bodies are brought in thermal contact they exchange heat irrespective of their temperatures. For example consider two bodies A and B which are in thermal contact. According to this law the body A which is at higher temperature will lose more heat and give it to B. Whereas the body B which is at lower temperature will lose less heat and give it to A. This means A will lose more heat and gainless whereas B will gain more heat and lose less. Thus there is a net loss of heat from A and net gain by B.

If there is no dissipation of heat and also no gain of heat from the surroundings that is if A and B form an isolated system in which there is heat interaction only between A and B, then according to the law of conservation of energy

$$\text{Heat lost by A} = \text{Heat gained by B}$$

$$(\Delta Q)_{lost} = (\Delta Q)_{gain}$$

$$C_A\, m_A\, \Delta T_A = C_B\, m_B\, \Delta T_B$$

Where

C_A = specific heat of material A

m_A = mass of A

ΔT_A = fall of temperature of A

The factors C_B and Cm have similar meaning.

However

ΔT_B = rise of temperature of B

A table of specific heat capacities is given below:

Table 11.3

Substance	Specific heat $Jkg^{-1} K^{-1}$	Substance	Specific heat $Jkg^{-1} K^{-1}$
Water	4200	Tungsten	135
Mercury	140	Silver	230
Ethyl alcohol	2430	Copper	390
Methylated sprit	250	Carbon	.502
Brass	380	Iron	450
Lead	130	Aluminium	900

11.22 MEASUREMENT OF SPECIFIC HEAT CAPACITY:

The branch of heat which is concerned with the measurement of heat is called Calorimeter. A calorimeter consists of a copper vessel which is enclosed inside a wooden box. The space between these two is filled with insulating materials like.paper, straw, wooden fittings, rubber, foam etc. The calorimeter is so designed as to minimise all types of heat losses. The top of the calorimeter is closed by a wooden lid which consists of two holes through which a thermometer and a stirrer are allowed to pass tightly.

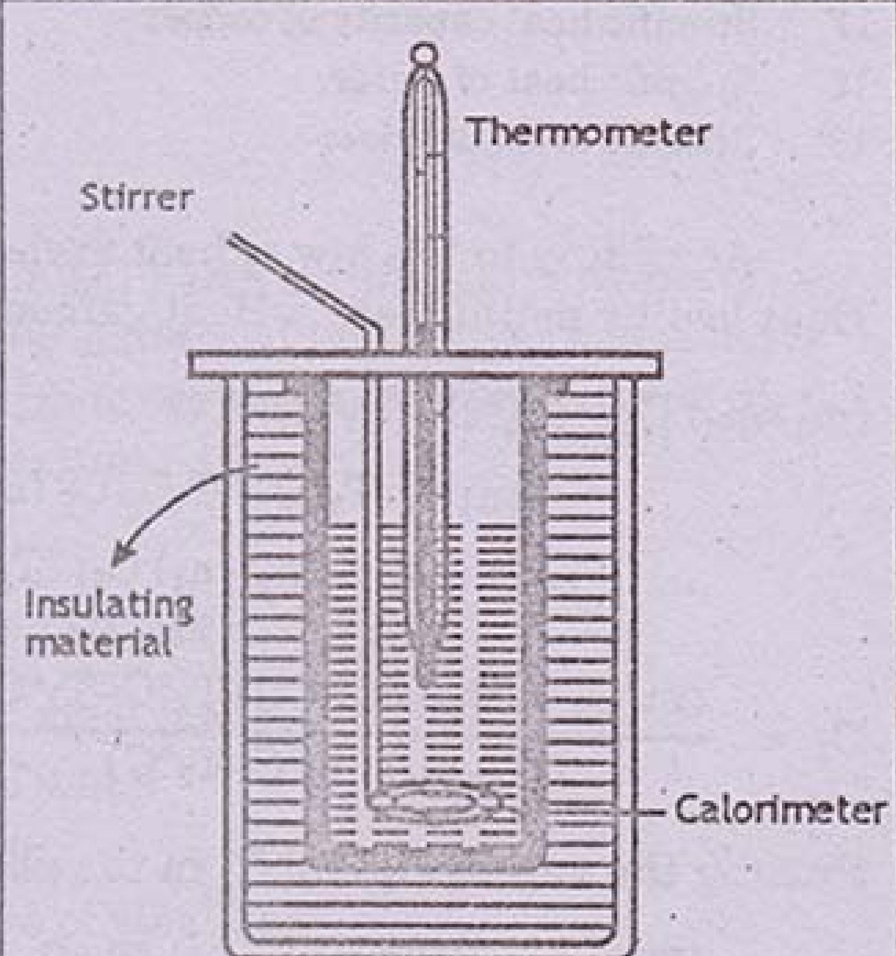


Fig. 11.12

To measure the specific heat capacity of a solid substance like a metal piece the following procedure is adopted.

First of all the mass of the calorimeter with stirrer is measured

with a physical balance. After that nearly two-thirds $\left(\frac{2}{3}\right)$ of the calorimeter is filled with water. The mass of the calorimeter with water is measured. The room temperature of the calorimeter with water is found by the thermometer. The metal piece is heated in a steam bath called "hypsometer" till it attains the temperature of the steam (generated at atmospheric pressure). It is then gently dropped into the water contained in the calorimeter which is now covered with the lid. The temperature of water begins to rise. The water is stirred so as to keep its temperature uniform throughout. The process of stirring is continued till the thermometer fitted in the lid of the calorimeter attains a final steady temperature. This temperature is noted. The calorimeter with water and metal piece (forming a mixture) is again weighed.

Calculations:

1.	Mass of the calorimeter	=	m_1 kg
2.	Mass of the calorimeter + water	=	m_2 kg
3.	Mass of calorimeter + water + metal piece	=	m_3 kg
4.	Mass of water	=	$(m_2 - m_1)$ kg
5.	Mass of the metal piece	=	$(m_3 - m_2)$ kg
6.	Room temperature of water	=	T_1°C
7.	Temperature of hot metal piece	=	T_2°C
8.	Final temperature of the mixture	=	T_3°C
9.	Rise in temperature of water	=	$T_3°C - T_1°C$
10.	Fall in temperature of metal piece	=	$T_2°C - T_3°C$
11.	Specific heat capacity of copper	=	$C_1 = 390\ Jkg^{-1}\ K^{-1}$
12.	Specific heat of water	=	$C_2 = 4200\ Jkg^{-1}\ K^{-1}$
13.	Specific of metal piece	=	C_3

According to the law of heat exchange:

Heat lost by metal piece = Heat gained by calorimeter+Water

$$(m_3-m_2)\,C_3 \times (T_2°C - T_3°C)$$

$$= m_1C_1\,(T_3°C - T_1°C) + (m_2-m_1)\,C_2\,(T_3°C-T_1°C)$$

$$= [m_1C_1 + (m_2-m_1)\,C_2]\,(T_3°C-T_1°C)$$

$$C_3 = \frac{[m_1C_1 + (m_2-m_1)\,C_2]\,(T_3°C-T_1°C)}{(m_2-m_1)\,(T_2°C-T_3°C)} \qquad (11.17)$$

Putting the values C_1 and C_2 in the above equation, we get

$$C_3 = \frac{[390m_1 + 4200\,(m_2-m_1)]\,(T_3°C-T_1°C)}{(m_2-m_1)\,(T_2°C-T_3°C)} \qquad (11.18)$$

Example 11.6: A piece of metal whose mass is 50 gram is heated so as to raise its temperature to 100°C. It is then dropped into water of mass 400 gram contained in a copper calorimeter whose mass is 400 gram at 20°C. If the final temperature of the mixture is 22.4°C, find the specific heat of the metal piece. Take $386 Jkg^{-1}k^{-1}$ as the specific heat of copper.

Solution:

m_1 = 50gm = 0.05 kg = mass of metal piece
m_2 = 400 gm = 0.4 kg = mass of water
m_3 = 400 gm = 0.4 kg = mass of calorimeter
T_1 = 100°C = temperature of hot metal piece
T_2 = 20°C = initial temperature of water
T_3 = 22.4°C = final temperature for mixture
C_1 = ? specific heat of metal piece
C_2 = $4200\ Jkg^{-1}K^{-1}$ = specific heat of water
C_3 = $386\ Jkg^{-1}K^{-1}$ = specific heat of copper

From the law of heat exchange
Heat lost by metal = Heat gained by water + Heat gained by calorimeter

$$0.05 \times C_1 (100°C-22.4°C) = 4200 \times 0.4 (22.4°C-20°C) + 386 \times 0.4 (22.4°C-20°C)$$

$$3.88\ C_1 = 4032 + 370.56 = 4402.56$$

$$C_1 = \frac{4402.56}{3.88} = 1134.68\ Jkg^{-1}K^{-1}$$

Example 11.7: A steel ball of mass m_1 = 80g has an initial temperature T_1 = 200°C. It is immersed in water of m_2 = 250g contained in a copper calorimeter of mass m_3 = 100g. The initial temperature of water and calorimeter is T_2 = 20°C. Find the final temperature of the mixture.

Solution:

Let T be the final temperature of the mixture.

m_1 = 80gm = 0.08 kg = mass of steel ball
m_2 = 250 gm = 0.25 kg = mass of water
m_3 = 100 gm = mass of calorimeter
T_1 = 200°C = initial temperature of steel ball
T_2 = 20°C = initial temperature of water and calorimeter

$C_1 = 450\ Jkg^{-1}K^{-1}$ = specific heat capacity of steel
$C_2 = 4200\ Jkg^{-1}K^{-1}$ = specific heat capacity of water
$C_3 = 385\ Jkg^{-1}K^{-1}$ = specific heat capacity of copper

$\Delta Q_1 = m_1C_1(T_1-T)$ = heat lost by steel ball
$= 0.08 \times 450\ (200-T) = 7200 - 36T$

$\Delta Q_2 = m_2C_2(T-T_2)$ = heat gained by water
$= 0.25 \times 4200\ (T-20) = 1050T - 21000$

$\Delta Q_3 = m_3C_3(T-20)$ = heat gained by calorimeter
$= 0.1 \times 385\ (T-20) = 38.5T - 770$

According to the law of heat exchange

Heat lost = Heat gained (total)

$$\Delta Q_1 = \Delta Q_2 + \Delta Q_3$$

$$7200 - 36T = 1050T - 21000 + 38.5T - 770$$

$$1050T + 36T + 38.5T = 21000 + 770 + 7200$$

$$1124.5T = 28970$$

$$T = \frac{28970}{1124.5} = 25.76K$$

Example 11.8: The temperature of a certain metal is raised to 100°C and then it is dropped in 200g of water at 15°C. If the temperature of the mixture at equilibrium is 21°C, calculate the specific heat capacity of the metal.

Solution:

Let C_1 be the specific heat capacity of the metal.

$C_2 = 4200\ Jkg^{-1}K^{-1}$ = specific heat capacity of water
m_1 = 500 g = 0.5 kg = mass of the metal piece
m_2 = 200 g = 0.2 kg = mass of water
T_1 = 100°C = initial temperature of the metal piece
T_2 = 15°C = initial temperature of water
T = 21°C = initial temperature of mixture

$\Delta Q_1 = m_1C_1(T_1-T)$ = heat lost by metal piece
$= 0.5 \times C_1\ (100-21) = 39.5C_1$

$\Delta Q_2 = m_2C_2(T-T_2)$ = heat gained by water
$= 0.2 \times 4200\ (21-15) = 0.2 \times 4200 \times 6 = 5040\ J$

According to the law of heat exchange

Heat lost = Heat gained (total)

$$39.5\ C_1 = 5040 \Rightarrow C_1 = \frac{5040}{39.5} = 127.59 Jkg^{-1}K^{-1}$$

11.23 SPECIFIC HEAT CAPACITY OF A LIQUID:

The method adopted to determine the specific heat capacity of a solid can be used for the determination of specific heat capacity of a liquid. Here the calorimeter is filled partially with the given liquid instead of water. A metal piece of specific heat capacity is heated in a hypsometer to a constant temperature and then dropped into the calorimeter containing the liquid. In this case the hot metal pieces loses heat whereas the liquid and the calorimeter gain heat. Using the law of heat exchange, the specific heat capacity of the liquid can be calculated.

11.24 LATENT HEAT:

In general when heat is added to a body, its temperature rises. For example when water is heated, its temperature rises. If we place a bowl of water in the sun for some time, its temperature rises because it absorbs heat-radian from the sun. But in certain situations even with the absorption of heat, temperature does not rise. This happens when the body changes its phase (from solid to liquid or from liquid to gas etc). During the change of phase (or state) the temperature of the body remains constant. With reference to the constancy of temperature, the added heat is called the latent heat or hidden heat (latent means hidden).

All solid substances require latent heat for fusing or melting. The latent heat of fusion depends on the nature of the substance. Here we give a table of melting and boiling points of certain substances.

Table 11.4

Substance	Normal Melting Point	Normal Boiling Point
Mercury	– 39°C	357°C
Sulphur	119°C	444.60°C
Lead	327.3°C	1750.0°C
Antimony	630.5°C	1440.0°C
Silver	960.8°C	2193°C
Gold	1063°C	2660.0°C
Copper	1083°C	1187.0°C
Hydrogen	– 259.31°C	– 252.89°C
Nitrogen	– 209.00°C	– 195.8°C
Oxygen	– 218.79°C	– 183.0°C

We now give in short some laws of fusion.

11.25 LAWS OF FUSION:

(i) Every substance changes its state from solid to liquid at a particular temperature (at normal pressure).

(ii) During the change of state the temperature remains constant.

(iii) One kilogram of every solid substance needs a definite quantity of heat energy to change its state from solid to liquid. It is called the latent heat of fusion of the substance.

(iv) Mostly substances show an increase in their volumes on melting (for example wax, ghee) while a few substances show decrease in their volumes on melting (for example ice).

(v) Melting points of those substances which show decrease in their volumes on melting, are lowered with the increase of pressure whereas melting points of those substance which show an increase their volumes are increased with increase of pressure.

We now consider two cases referring to latent heal. They are (a) Latent heat of fusion of ice; and (b) Latent heat of vaporization of water.

(a) Latent heat of fusion of ice:

Definition: It is the amount of heat energy needed to melt 1kg of ice at 0°C completely to 1kg of water at 0°C. The experimental values of the latent heat of fusion of ice has been found to be 3.36×10^5 J/kg. This also means that if 1kg of water at 0°C is completely frozen to ice, the same amount of heat energy will be given off from it.

If ΔQ is the amount of heat energy needed to melt mkg of ice at 0°C completely to mkg of water at 0°C, then

$$\Delta Q = mL \Rightarrow L = \frac{\Delta Q}{m} \quad \text{.....................} \quad (11.19)$$

where L is the latent heat of fusion of ice.

Explanation:

If we lake a piece of ice at 0°C and heat it, its temperature does not rise until the whole of the ice has been melted to water at the same temperature (0°C). Here the heat energy added is used up in loosening the bonds between the molecules. This helps in overcoming the force of attraction between the molecules. The result is that the molecules begin to vibrate vigorously. The vibrational amplitude of the molecules become so large that the bands them break and the molecules become free. These molecules thus form water in which they move about freely.

11.26 VAPORIZATION AND LATENT HEAT OF VAPORIZATION:

Vaporization is the process in which a substance changes its state from liquid to gaseous. Thus in the process of vaporization water changes from liquid state to steam.

When water at 0°C is heated, its temperature can rise up to 100°C which is the boiling point of pure water at atmospheric pressure (boiling point increases with the increase of pressure). Further heating does not cause any change on the boiling point. By absorbing heat the boiling water changes its state from liquid to gaseous (vapour, that is it transforms into steam). During the formation of steam, the temperature remains constant (100°C) till the whole of water is converted into steam. The heat added is therefore termed as latent or hidden. We now define latent heat of vaporization of water. The amount of heat required to change 1kg of water at its boiling point (100°C) to steam at the same temperature. The experimental value of latent heat of vaporization of water is found to be 2260000 J/kg (2.26×10^6 J/kg). The same amount of heat energy is given off when steam condenses to water at 100°C.

Explanation: The latent heat of vaporization is used up to separate the close liquid molecules. Work is required to overcome the strong inter-molecular forces of attraction of the liquid (water) molecules. This is done by the latent heat of vaporization. Thus the molecules become free and form the gaseous state.

Example 11.9

Heat is supplied at a constant rate to 0.4kg of ice at 0°C in a copper container. All the ice is converted into water in 5 minutes. How much time will be required (i) to raise the temperature of water to its boiling point (100°C) and (ii) to change all the water at 100°C into steam.

Solution:

$m = 0.4\text{kg}$ = mas of ice

$L = 3.36 \times 10^5$ J/kg = Latent heat of fusion of ice.

$t = 5$ minutes $= 5 \times 60$ sec $= 300$ sec = time taken

ΔQ_1 = heat absorbed by ice

$= m \times L = 0.4 \times 3.36 \times 10^5 = 1.344 \times 10^5 J$

$$\text{Rate of supply of heat} = \frac{\Delta Q_1}{t} = \frac{1.344 \times 10^5}{300} = 0.448 \times 10^3$$

$= 448$ J/sec.

(i) ΔQ_1 = Heat required to raise the temp of 0.4 kg of water from 0°C to 100°C $= 0.4 \times 4200 \times 100 = 168000$ J

$$\text{time taken} = \frac{168000}{448} = 6 \text{ min-15 sec.}$$

(ii) ΔQ_3 = Heat required to convert 0.4kg of water at 100°C into steamat 100°C.

$$= mL = 0.4 \times 2.26 \times 10^6 J = 0.904 \times 10^6 J$$

$$\text{time taken} = \frac{\Delta Q_3}{448} = \frac{0.904 \times 10^6}{448} \text{ sec} = 2.0178 \times 10^3$$

$$= 2018 \text{ sec} = \frac{2018}{60} \text{ min} = 33 \text{ min-38 sec.}$$

$$\text{total time taken} = 375 \text{ sec} + 2018 \text{ sec.}$$

$$= 2393 \text{ sec} \approx 40 \text{ min.}$$

Example 11.10 An electric heater rod is dipped in a vessel containing water at 0°C. The electric rod produces heat at the rate of 3000J/s. The vessel with its contents is maintained at 0°C by adding ice at the rate of 9g/s. Calculate the latent heat of fusion of ice.

If the total heat capacity of water, the vessel and its contents is 12000 J/k, at what rate does the temperature start rising when the supply of ice is stopped?

Solution:

(i) ΔQ = 3000 J/s = heat produced by the electric heater.
m = 9g/s = 9×10^{-3} kg/s = mass of ice added per second.
L = ? = latent heat of fusion of ice.

Heat absorbed by ice/sec = Heat produced by heater/sec

$$mL = 3000 \Rightarrow L = \frac{3000}{m} = \frac{3000}{9 \times 10^{-3}} = 3.33 \times 10^5 \text{ J/kg}$$

(ii) Let ΔT be the rate of rise of temperature when the supply of ice is cut off.

C = 12000 J/K = heat capacity
ΔQ = 300 J/sec.
$\Delta Q = C \times \Delta T$

$$\Delta T = \frac{\Delta Q}{C} = \frac{3000 \text{ J/s}}{12000 \text{ J/k}} = 0.25 \text{ k/s}$$

11.27 TRANSMISSION OF HEAT:

If we immerse our hands in a warm water, we feel hot. Here heat is transferred from warm water to our body. Similarly if we place our hands in ice cold water, we feel cold because here heat is transferred from our body to the cold water. We get heat energy from the sun due to transfer of heat.

Thus heat travels from one point to another or from one place to another because of difference of temperature. There are three different modes of transfer of heat. They are

(i) Conduction, (ii) Convection; and (iii) Radiation

(i) **Conduction:**

This is the process in which heat travel from particles to particles (microscopically from atoms to atoms or from molecules to molecules) in a body. Without any visible motion of any part of the body. This mode of transfer of heat takes place mainly in solids where the atoms (or molecules) are very closely situated. When a body is heated its temperature rises at the heating place. Due to the rise in temperature, the average kinetic energy of the atoms (or molecules) increases. Hence the atoms begin to vibrate with greater and greater amplitude with the rise of temperature about their mean positions. This results in the collision of atoms and the heat absorbed by an atom is transferred to the neighbouring atoms through collision. Thus heat is carried through the body from atoms to atoms. This can be demonstrated by a simple experiment. A long metal bar is covered with a thin layer of vax at one end. The vax coated end is heated by placing it under a flame. This end absorbs heat energy and as a result the vax begins to melt. As the time passes on, the melting starting at the hot end goes further and further along the vax coated length of the rod. This shows that the heat is carried continuously from the hot end towards the cold end of the rod.

The rate of transmission of heat depends on the nature of the substance. Some substances are quickly heated by this process. They are all metallic solids except mercury which is liquid. They are therefore called good conductors of heat. In these substances the heat conduction is greatly aided by the presence of free electrons which move through out the body of the metal and carry heat energy with them.

There is another class of substances which are not conduct heat easily from one point of the body to another. They are poor conductors of heat. Glass, wood, asbestos, rubber and plastic are a few examples of poor conductors of heat. A wooden stick can burn at one end leaving the other end relatively cold. These substances are, sometimes, also called bad conductors or thermal

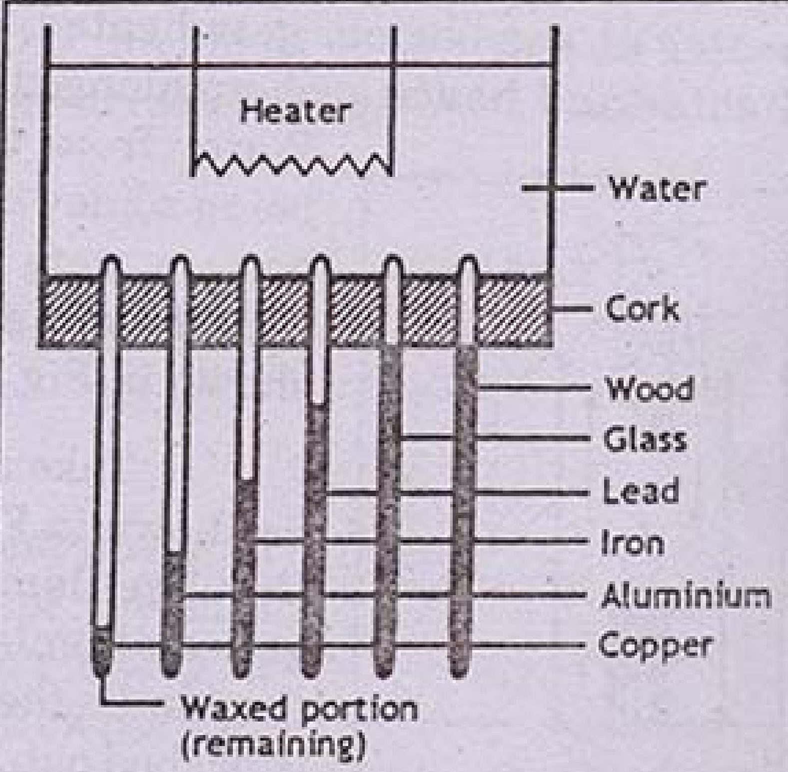


Fig. 11.13: Solid Heat Conductors

insulators. Here we have experimental set up showing the ability of different substances to conduct. Here water is heated in a container which has various types of rods, initially coated with a thin layer of vax. The mode of heat transfer is mainly conduction, downwards along the rods. The melted length of the coated vax along a rod gives its ability to conduct. As clear from the Fig. (11.13) the melted portion of vax is maximum along copper rod and minimum along the wooden and glass rod, which are good insulators of heat. Copper is the best conductor of heat.

(ii) Convection:

It is the mode of transfer of heat through a fluid (that is through a liquid or a gas) by the actual movement of the fluid (microscopically the molecules of the fluid). The fluid receives heat directly from the source and gets heated. it expands, becomes lighter and therefore rises up. Its place is the taken by somewhat cooler and denser fluid. The circulation of fluid sets up convection current. The same process holds in the boiling of water which is taken in an electric kettle. The heater of the kettle is normally placed near the bottom of the kettle so that as the water lying at the bottom of the kettle is heated, its expands and gets lighter. Being lighter, it rises up the cooler section of water being denser moves down and heated. This process is repeated due to convection currents set up until the whole of water reaches the boiling point.

In the following examples we will demonstrate how the transfer of heat takes place in convection.

1. Take a flask containing water. Now add a large crystal of $KMnO_4$ to the water. The flask is heated, coloured streaks of water rise up. It is because of the fact that the water at the bottom gets heated, expands and became lighter and hence goes up along this sides of the vessel. Water from the sides of the flask being somewhat denser reaches the bottom, gets heated and rises up, thus forming coloured streaks as shown in Fig. (11.14).

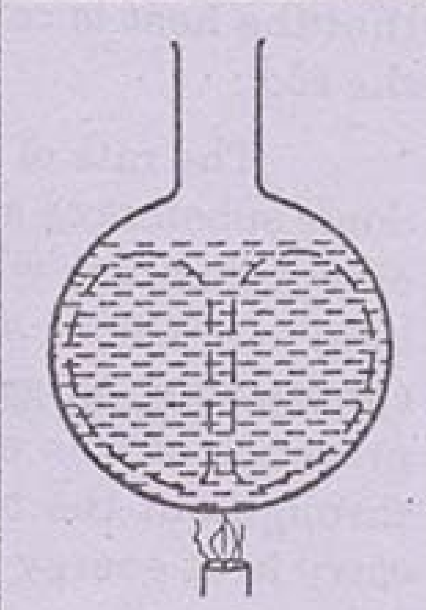


Fig. 11.14

2. Take a candle and fix it at the bottom of a cylinder as shown in Fig. (11.15). Light the candle. It will be found that the flame becomes weaker and weaker and finally gets extinguished. It is due to the fact that by burning of the candle the air in the cylinder gets heated, expands and is pushed out. There is no fresh supply of air for the burning of the candle.

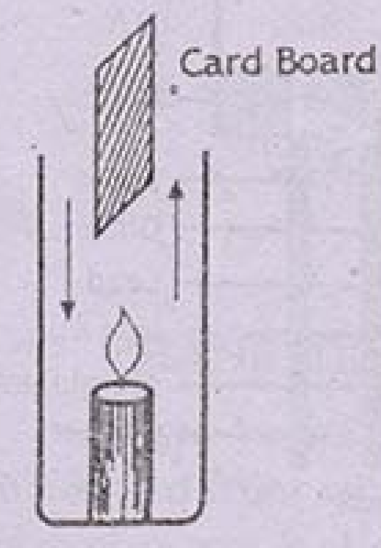


Fig. 11.15

Now take a cardboard and hold it inside the cylinder dividing the space above the candle into two parts. Again light up the candle. It will continue to burn. Here, the air above the flame gets heated and goes upthrough one of the sides of the card board. The fresh air moves in through the other side, forming convection currents and so the candle continues to burn.

(iii) Radiation:

In the process of radiation not only heat energy is transferred or carried. All objects emit energy at all temperatures from their surfaces. The energy emitted and carried through radiation is called radient energy. The radient energy is carried by electromagnetic-waves.

A hot piece of metal gives off light. Its colour depends on the temperature of the metal, going from red to yellow to white as it becomes hotter and hotter. The light emitted corresponding to different colours is a part of electromagnetic waves. An object need not be hot to emit radient energy. At room temperature most of the radiation is found in infrared region.

All the energies emitted and carried by radiation are electromagnetic in nature. Light of every colour (from infrared to ultraviolet), radio and TV waves, microwaves and x-rays are all electromagnetic waves. The difference lies in their frequencies or wave length.

11.28 THERMAL CONDUCTIVITY:

We have discussed conduction of heat in detail. We now study the ability of a substance to conduct. This ability is the measure of thermal conductivity of a substance, which is a thermal property of a substance.

Consider a slab of thickness ΔL with opposite faces each of area A maintained at temperatures T_1 and T_2 so that $T_1 > T_2$. Heat is conducted in the direction of fall of temperature across the thickness of the slab. It is found experimentally that the heat ΔQ conducted across the two faces varies as

$$\Delta Q \propto \frac{A(T_1 - T_2)\,\Delta t}{\Delta L}$$

$$\Delta Q = K\,\frac{A\,.\,\Delta T\,.\,\Delta t}{\Delta L} \quad \text{.....................} \quad (11.20)$$

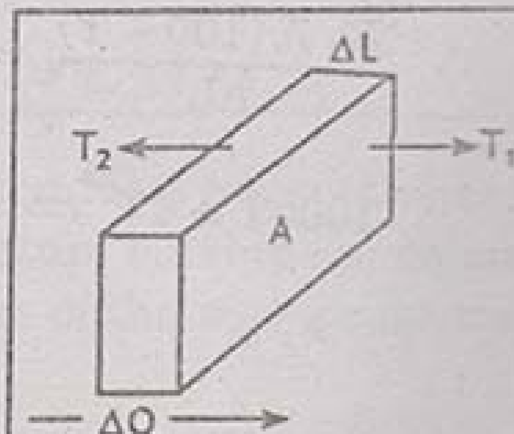


Fig. 11.16

Where Δt is the time internal of heat flow and 'K' is a constant called the coefficient of thermal conductivity. Thus the coefficient conductivity is numerically equal the amount of heat flowing per second across the opposite faces of a cube of unit size whose faces are maintained at a difference of temperature equal to 1°C (or 1k). It depends on the nature of

the substance. It is large for metals and small for non-metallic solids, liquids and gases. In the system the unit of thermal conductivity is $Jm^{-1} K^{-1} S^{-1}$.

Example 11.11 Assume that the thermal conductivity of copper is 4 times the thermal conductivity of brass. Two rods of copper and brass having the same length and same cross-section are joined end to end. The free end of the copper rod is kept at 0°C and the free end of brass rod is kept at 100°C. Calculate the temperature of the junction of the two rods. Neglect radiation losses.

Solution:

Let T be the temperature of the junction.

K_1 = thermal conductivity of brass
K_2 = $4K_1$ = Conductivity of copper
A = area of cross section of each rod.

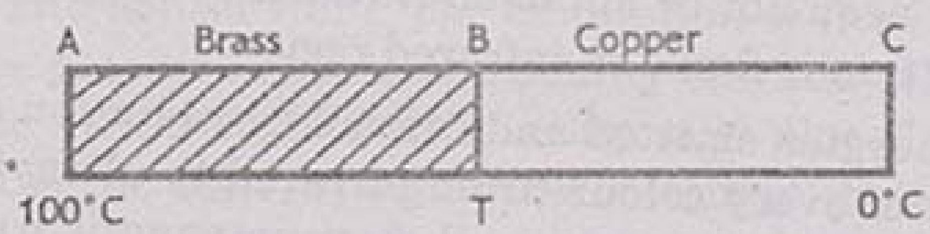


Fig. 11.17

In equilibrium

$$\frac{K_1 A_1 (T_1 - T)}{\Delta L_1} = \frac{K_2 A_2 (T - 0)}{\Delta L_2}$$

Here $\Delta L_1 = \Delta L_2 = \Delta L$

$A_1 = A_2 = A$

$K_2 = 4K_1$

$$\therefore \quad \frac{K_1 A (100 - T)}{\Delta L} = \frac{4K_1 (T - 0)}{\Delta L}$$

$$100 - T = 4T \Rightarrow 5T = 100 \Rightarrow T = \frac{\cancel{100}^{\,20}}{5} = 20°C$$

$$T = 20°C$$

11.29 PRACTICAL APPLICATION OF CONDUCTION OF HEAT:

(1) Ice box has a double wall, made of tin or iron. The space between the two walls is filled with cork or felt which are poor conductors of heat. They prevent the flow of out side heat into the box, thus keeping the ice from melting.

(2) Woolen clothes have fine pores filled with air. Air and wool are bad conductors of heat. Thus the heat from the body does not flow out to the atmosphere. Thus the woolen clothes keep the body warm in winter.

(3) In cold countries windows are provided with double doors. The air in the space between the two doors forms a non conducting layer and so heat cannot flow out from inside the room.

In hot countries as well, the windows are made with double doors. Heat can not flow in from outside because of the presence of air between the two doors.

(4) When a stopper, fitted tightly to the bottle is to be remained, the neck of the bottle is gently heated. It expands slightly on heating. Since glass is a bad conductor of heat, the heat does not reach the stopper. Thus it can be removed easily.

(5) Davy's Safety Lamp: It is one of the most important applications of conduction of heat. The principle of Davy's safety lamp can be understood from a simple example in which a wire gauze is placed over a bunsen burner. The gas coming from the burner is lit above the wire gauze as shown in Fig. (11.18a). A flame appears at the top surface of the wire gauze. The gas coming out from the burner below the wire gauze does not get sufficiently heat for ignition. The reason is that the wire gauze conducts away the heat of the flame above it and so the temperature at the lower surface of the gauze does not reach the ignition point.

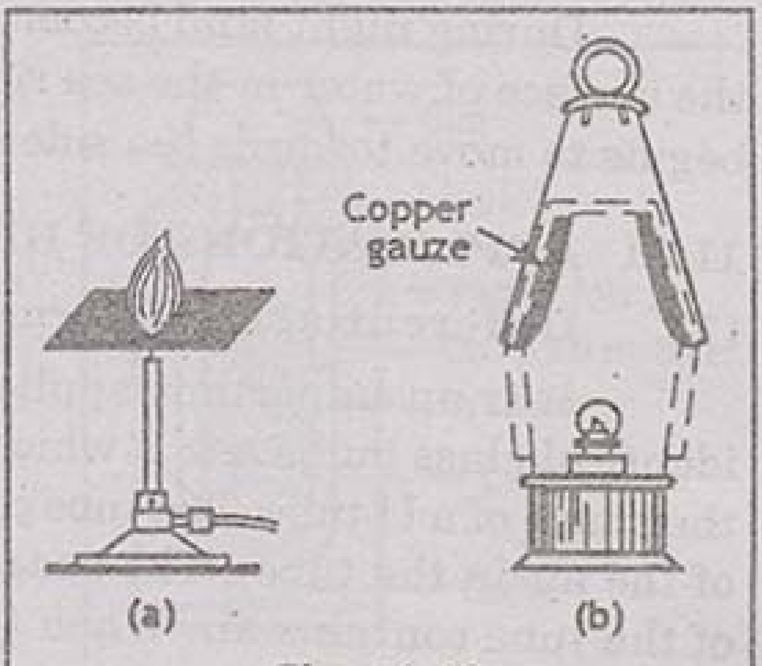


Fig. 11.18

In Davy's safety lamp, a cylindrical metal gauze of high thermal conductivity surround the flame as shown in Fig. (11.18b). When this lamp is taken inside a mine, the explosive gases present in the mine are not ignited because the wire gauze in the form a cylinder conducts away the heat of the flame of the lamp. The result is that the temperature outside the gauze remains below the ignition point of the gases. In absence of the wire gauze the gases outside can explode.

11.30 PRACTICAL APPLICATIONS OF CONVECTION OF HEAT:

Here we give a few applications of convection of heat.

(1) **Ventilation:** From health point of view every lining room of a building should be provided with ventilators near the ceiling. Due to respiration of the

persons sitting or sleeping in the room the air in the room gets warmer and hence is less dense. It rises up and goes out side through the ventilator. Fresh air comes in the room through the doors and windows. Thus a convection current of air is maintain.

2. **Trade wings:** At the equator the surface of the earth gets heated more than at the poles. This results in the movement of the warm air from the equator to the poles of the cold air moves towards the equator. Because of the rotation of the earth (from west to east) the air in the Northern Hemisphere seems to be coming from north-east instead from north. In the South Hemisphere the air from the south hole appear to be coming from south west. These winds are called trade winds because in olden days these winds were used by traders for sailing their ships.

3. **Land and sea breezes:** Land is a better conductor of heat than water. Hence in day-time the land gets hotter than water in the sea. The air above the land becomes warm and rises up being lighter and some what cold air above sea surface moves towards the sea short. This is known as seabreeze. Thus convection currents fair our set-up.

During night land becomes cooler than water and so the warm air over the surface of water in the sea rises up. The air on the land near the sea shore begins to move towards sea side and is called land breeze.

11.31 APPLICATIONS OF HEAT RADIATION:

(1) Differential Air Thermoscope:

It is an important application of radiation of heat. It consists of two identical glass bulbs A & B which are connected by a narrow glass tubing have the shape of a U-tube. The tube consists of sulfuric acid (so as to absorb moisture of the air in the tube). The space above the levels of the acid in the two arms of the tube contains air. When the bulbs are at the same temperature there is no difference in the level of the acid in the two limbs. The bulb A is coated with lamp black so that it may completely absorb the heat radiation falling on it.

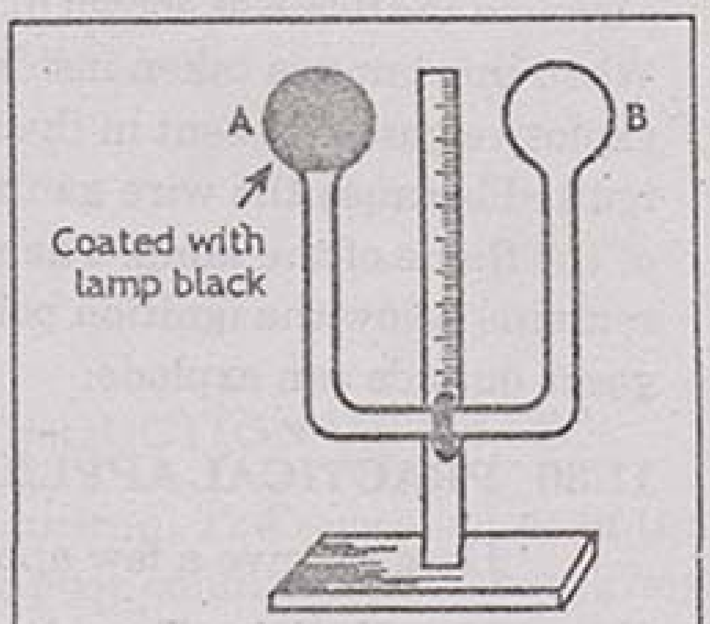


Fig. 11.19

Now the bulb "A" is exposed to heat radiation. It absorbs whole of the radiation falling on it. As a result the air in the bulb A gets heated, expands and presses down the acid in the limb. Thus we have a difference in the level of the liquid in the two limbs. This thermoscope is very sensitive and can detect

radiation of very weak intensity for example the radiation coming from a distant candle.

(2) **Boy's Radiomicrometer:**

It is a very sensitive device and can detect heat radiation of very weak intensity from example radiation coming from a distant candle can be detected. It is a combination of a moving coil galvanometer and a thermocouple. It consists of a single loop of silver or copper wire A. The lower ends of the wire are soldered to a copper disc which is coated with lamp black. The disc is exposed to heat radiation and as a result thermo-electric current of produced in the couple made of hismuth and antimony, say, begins to flow in the wire A. Hence we get a current in the galvanometer. The deflection produced in the galvanometer can be measured by using lamp and scale arrangement.

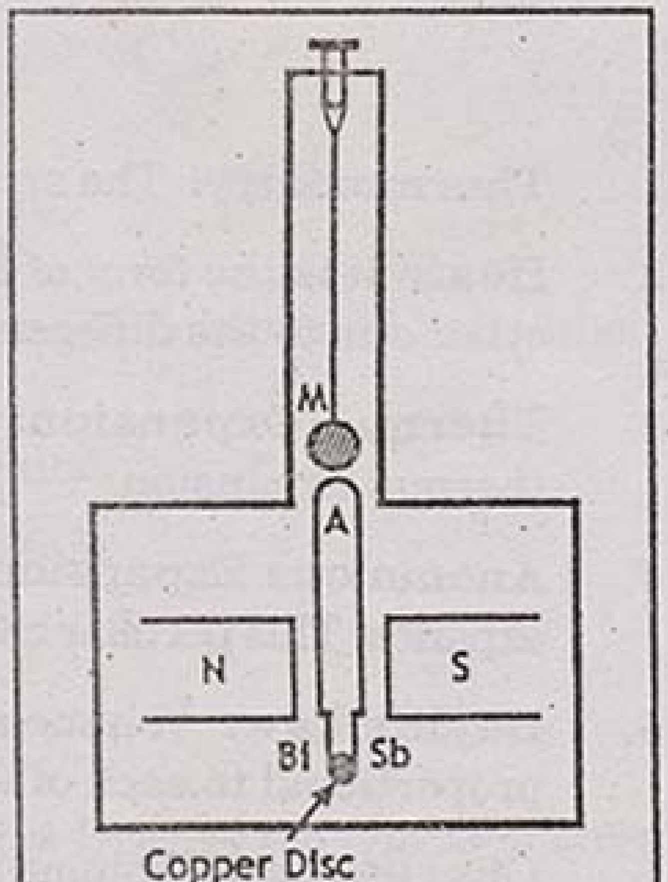


Fig. 11.20: Boy's Radiomicrometer to measure the intensity of heat radion

11.32 THERMOS FLASK:

A thermos flask is a device where all the three modes of transfer of heat are applied. It consists of a double walled glass bottle. The inner surface of the outer wall and the outer surface of the inner wall are highly polished. The space between the walls is evacuated and scaled. The whole system is enclosed with in a metal case which is provided with a cork at the bottom and a pad of felt at the neck for safety as shown in Fig. (11.21). Glass is a poor conductor of heat where as air, cork, felt etc. which are used in thermos flask are all had conductors of heat. Hence they prevent any loss of due to conduction.

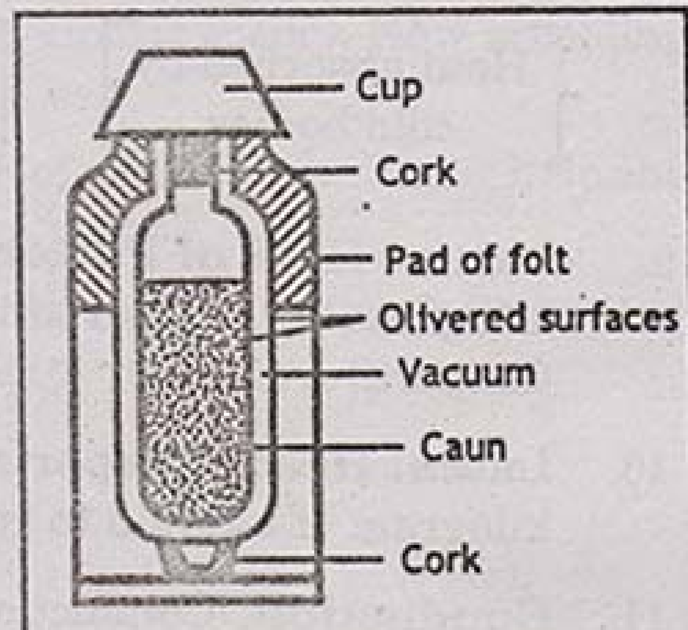


Fig. 11.21: Thermos Flask

When a hot liquid is kept in the bottle, it remains hot for a long time. Any heat radiation coming from the hot liquid is reflected back from the inner surface of the outer wall. Also the heat from the liquid can not flow out through conduction or convection because of the empty space between the walls. Thus all the three processes of heat-transfer fail to transfer heat.

Any heat energy through radiation conduction, convection or coming from outside can not reach the liquid because of the highly polished surfaces and the empty space between them.

SUMMARY

1. **Thermometry:** The art of measuring temperature is called thermometry.
2. **Heat:** It is the form of energy which is transferred from one body to the other due to the difference in temperature.
3. **Thermal Expansion:** Expansion of substances on heating is called thermal expansion.
4. **Anomalous Expansion of Water:** Water contracts from 0°C to 4°C then expands. This peculiar characteristic of water is called anomalous expansion.
5. **Boyle's Law:** Volume and pressure of a gas of fixed mass are inversely proportional to each other provided the temperature is kept constant.
6. **Charle's Law:** Volume and temperature of a gas of fixed mass are directly proportional to each other provided the pressure of the gas is kept constant.
7. **Specific Heat:** the quantity of heat required to change temperature of one kg mass of a substance by one kelvin is called specific heat.
8. **Law of Heat Exchange:**

$$\begin{bmatrix}\text{Heat absorbed by} \\ \text{cold bodies}\end{bmatrix} = \begin{bmatrix}\text{Heat released by} \\ \text{hot bodies}\end{bmatrix}$$

9. **Latent Heat of Fusion:** The quantity of heat required to change one kilogram of a solid substance to liquid state without changing the temperature.
10. **Latent Heat of Vaporization:** the amount of heat needed to change one kilogram of a liquid to vapour or gaseous state at a fixed temperature.
11. **Conduction:** Conduction is the process in which heat is transferred from one part of the body to the other by interaction of electrons and molecules.
12. **Convection:** transmission of heat due to actual movement of molecules of the substance from one place to the other.
13. **Radiation:** Radiation is the process of heat transmission in which heat energy is transferred from one place to the other in the forms of waves without affecting the medium.
14. **Thermal Conductivity:** When opposite faces of a metre cube of a body are kept at a temperature difference of 1K, then the quantity of heat reaching from one face to the other in one second is said to be coefficient of thermal conductivity of that body.

QUESTIONS:

11.1 Write answer to the question:

(i) Differentiate between heat and temperature.

(ii) Describe in detail the construction and working of Celsius and Fahrenheit scales of temperature.

(iii) What are gas laws? Derive general gas equation.

(iv) Define coefficient of linear expansion. Show that

$$\beta = 3\alpha$$

where

α = coefficient of linear expansion

and β = coefficient of volume expansion.

(v) Define heat capacity and specific heat capacity. Discuss the effects of large and small specific heat capacities.

(vi) Define latent heat. Describe a method to determine the latent heat of fusion of ice.

(vii) What are different modes of transfer of heat? Discuss convection and radiation in detail.

(viii) Explain the phenomen of evaporation using kinetic theory. How is cooling caused by evaporation.

(ix) What do you mean by anomalous expansion of water. How does it help the aquatic animals to save their lives in frozen seas?

(x) Explain why is it more dangerous to burn from steam than from boiling water although both are at the same temperature?

11.2 Fill in the blanks.

(i) 37°C = ______ °F.

(ii) 100°C = ______ K.

(iii) Water expands on cooling from 4°C to 0°C. This expanse of water is called ________.

(iv) The heat energy needed to vaporize 1kg of boiling water completely into steam is called ________.

(v) The law of heat exchange states that
Heat ________ = Heat ________

(vi) Specific heat capacity of water = ________.

(vii) Specific heat of water is ______ than the specific heat of copper.

(viii) In the system International the unit of specific heat is ________.

(ix) Latent heat of fusion of ice is ________.

(x) For an ideal gas Boyle's law holds at constant ________.

(xi) During the change of state ________ remains constant.

(xii) Melting point of ice ________ with increase of pressure.

(xiii) Boiling point of a liquid ________ with increase of pressure.

(xiv) Thermos that is a device which controls ________ in a certain range.

(xv) A ________ strip is used for the operation of a fire alarm.

11.3 Choose the best answers:

(i) The diagram shows a gas (vapour) being cooled; its temperature is plotted against time. Which portion of the graph represents the phase of pure liquid?

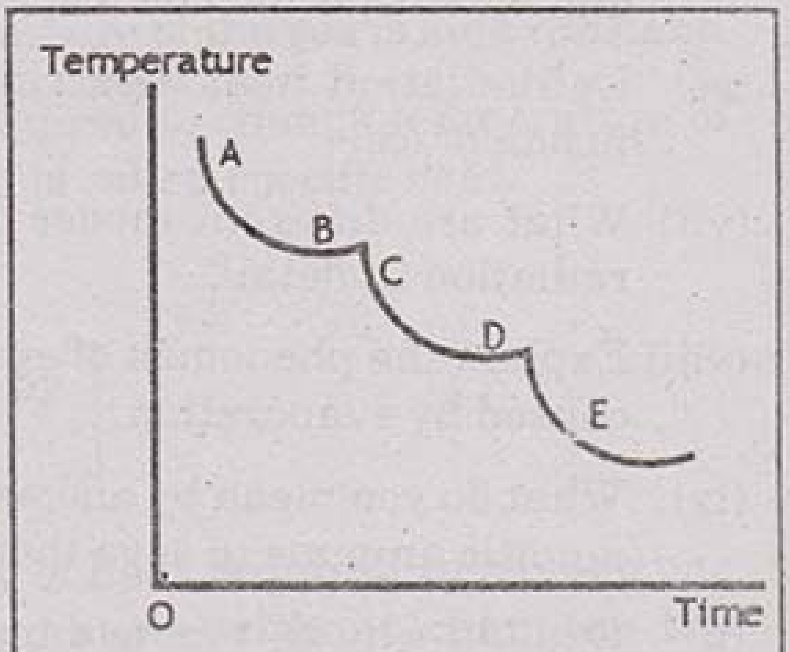


Fig. 11.22

(ii) The energy released in portion B of the graph (Fig. 11.22) is called a Latent heat of fusion

b Latent heat of vaporization

c Latent heat of combustion

d Latent heat of solidification

e Latent heat of sublimation.

(iii) A given volume of helium gas is heated by 20°C at constant external pressure. Its volume will

a. decrease b. increase c. remains the same

d. increase by 20% e. decrease by 20%.

(iv) The temperature of a substance changes from –20°C to 20°C. What is the temperature change in Kelvin's scale?

a. 0K b. 20K c. 40K

d. 293K e. 400K.

(v) At S.T.P. pure water boils at

a. 0K b. 100K c. 273K

d. 237K e. 227K.

(vi) When a piece of red-hot iron is thrust into boiling water, the temperature of the boiling water will

a. increase b. remain the same

c. decrease d. fluctuate

e. reach the temperature of the iron.

(vii) Mercury is commonly used as the working liquid in laboratory thermometer because it.

a. has a high freezing point

b. has a high density

c. is a liquid metal

d. is silvery

e. expands evenly and proportionally when heated.

(viii) How can you increase the sensitivity of a liquid-in-glass thermometer?

a. use a liquid which conducts heat better

b. use a thin-walled bulb

c. use a longer bulb

d. use a liquid of very high boiling point

e. use a tube, having a narrower bore.

(ix) The heat from an open fire in the kitchen reaches cook's face by

a. radiation only

b. convection only

c. conduction only

d. both radiation and convection

e. both radiation and conduction.

(x) Thermal expansion and contraction of metals are used in each of the following <u>except</u> .

a. riveting steel plates

b. fixing axles for wheels

c. pressure cooker

d. fire alarm.

(xi) bimetallic strip is used in temperature control devices because the two metals used in a bimetallic strip

a. expand unequally when heated

b. are good conductors of heat

c. conduct heat at different rates

d. can bend easily.

(xii) The reason for bursting water pipes during very cold weather is that

a. water pipe contract when cooled
b. water expands on freezing
c. ice expand on melting
d. the structure of the material of the pipes is weakened at low temperature.

(xiii) The volume of a gas when heated increases much more than for a solid or liquid because

a. the particles of a gas expand more
b. the attractive forces between the molecules of a gas are negligible
c. the particles of solids and liquids can not move
d. the molecules of a gas are lighter.

(xiv) The pressure of 30 litres of a gas at S.T.P. is increased to 3 atmospheres, keeping the temperature constant. What is the new volume?

a. 1 litre   b. 3 litres
c. 10 litres   d. 90 litres.

(xv) In which of these lists are the substances correctly arranged so that their densities are in the increasing order?

a. water at 4°C, ice at 0°C, water at 23°C
b. water at 0°C, water at 4°C, water at 23°C
c. water at 23°C, ice at 0°Cm water at 4°C
d. ice at 0°C, water at 23°C, water at 4°C.

(xvi) You are provided with two cups of hot tea you are to wait for a friend for short time. Would you

a. sugar first and wait for your friend?
b. wait for your friend and add sugar after the friend comes?

PROBLEMS

11.1 A metal rod of diameter 1cm measures 50cm in length at 20°C. When it is heated to 95°C, its length becomes 50.06cm. What is the coefficient of linear expansion of the rod? What will be its length when it is cooled to 0°C?

(Ans. $\alpha = 16 \times 10^{-6}/°C$, $L_0 = 49.984cm$)

11.2 The difference between the lengths of two rods A and B is 60cm at all temperatures. Find their original lengths at 0°C. Given $\alpha_A = 18 \times 10^{-6}/°C$, $\alpha_B = 27 \times 10^{-6}/°C$

[Ans. $(L_A)_0 = 180cm$, $(L_B)_0 = 120cm$]

11.3 WA gas of a given mass at a pressure of 50cm of Hg is heated from 27°C to 97°C. If the volume is maintained constant, calculate the pressure exerted by the gas.

(Ans. 61.67cm of Hg)

11.4 If a given mass of a gas has a volume of $4.5 \times 10^{-5} m^3$ at a pressure of 30.0 kP_a, what will be the volume of the gas if the pressure is increased to 50.0 kP_a while the temperature is kept constant?

(Ans. $2.7 \times 10^{-5} m^3$)

11.5 A bubble of air rises from the bottom of a pond to the surface. Just before the bubble reaches the surface it breaks and the volume was double its original volume. The water has a uniform temperature and its density is 1000 kg m^{-3}. Atmospheric pressure is 10^5 P_a. Estimate the depth of the pond.

(Ans. 10.2m)

11.6 An electric heater of power 500W raises the temperature of 4.0 kg of a liquid from 10°C to 15°C in 100S.

Calculate

(i) the heat capacity of the 4.0 kg of the liquid

(ii) the specific heat capacity of the liquid

(Ans. (i) 12000 J/K, (ii) 3000 J/Kg.K)

11.7 A 2KW steel kettle of mass 1kg contains 1.5kg of water at 30°C. What is the time taken to boil the water if the specific heat capacity of steel is 460J kg^{-1} $°C^{-1}$ and that of water is 4200J kg^{-1} K^{-1}?

Calculate

(i) the heat capacity of the 4.0 kg of the liquid

(ii) the specific heat capacity of the liquid

(Ans. 237 S)

11.8 1 litre of water at 100°C is added to 5 litres of water at 30°C. What is the final temperature of water?

(Ans. 41.7°C)

11.9 Two liquids A and B are at temperatures 60°C and 20°C respectively. Their masses are in the ratio 3:4 and their specific heats are in the ratio 4:5. Calculate the final temperature of the mixture if the liquids A and B are mixed. Neglect the water-equivalent of the calorimeter.

(Ans. 35°C)

11.10 1kg of water is contained in a 1.25 kilowatt kettle. Assuming that the heart capacity of the kettle is negligible, calculate the time taken for the temperature of water to rise from 25°C to its boiling point at 100°C.

(Ans. 4min-12sec)

11.11 20gram of ice at −10°C is converted into steam at 100°C. Find out the total heat energy required to accomplish the change. Given the specific heat of ice, latent heat of ice and latent heat of steam as 2.1 J/g, 326 J/g and 2268 J/g respectively.

(Ans. 60900 J)

11.12 Two vessels made of different metals are similar in shape and size. They are fully filled with ice at 0°C. By the heat from outside all the ice in one vessel melts in 25 minutes and that in the other vessel in 20 minutes. Compare their conductivities.

(Ans. $\frac{K_1}{K_2} = 0.8$)

CHAPTER - 12

WAVES AND SOUND

12.1 INTRODUCTION

You must have wondered how sound reaches our ears. Why do we fix an antenna on our roof top if we have a TV? How does light from the sun and other stars reach us? The answers to all these amazing questions lies in the study of *waves*. In this chapter we shall study the wave phenomena and their manifestation in sound.

Often children playing near a pond drop a pebble or stone into the pond water to produce circular shaped ripples on the surface of calm and still water. The children also amuse themselves when they give continual jerks to a rope and try to make a snake like motion of the rope. The circular shaped ripples formed on the surface of water and the snakelike motion of the rope are examples of motion which is known as wave motion. These are examples of visible waves. But there are other waves which are invisible. In fact we are surrounded by invisible waves. Sound, light and radio waves are example of these invisible waves. Whether visible or not, all waves have several common properties and therefore by studying visible waves we would understand the invisible waves in sound and light.

12.2 OSCILLATIONS

Any motion that repeats itself in equal intervals of time is called periodic motion. This periodic motion is often called harmonic motion.

If a particle in periodic motion moves back and forth over the same path, we call the motion oscillatory or vibratory. The world is full of oscillatory motions. Some common examples are, the swinging bob of a pendulum, balance wheel of a watch, a violin string and mass attached to a spring. Less obvious, but at least as important is the example of the oscillating air molecules carrying sound waves produced by the violin. This type of motion is quite different from the translatory and rotatory motion of the body.

The phenomena of elasticity is closely linked with that of vibration (or oscillation). It is a common experience that when some elastic material, for example, a mass m attached to one end of spring and placed on a friction less horizontal surface as shown in the Fig. 12.1 is displaced from its normal or equilibrium position and then released, it will move through the equilibrium position and become extended in the opposite direction. Having extended itself to the other side of its normal position, the motion is reversed and is directed back toward the equilibrium position where again it goes to other side as shown in Fig: 12.2.

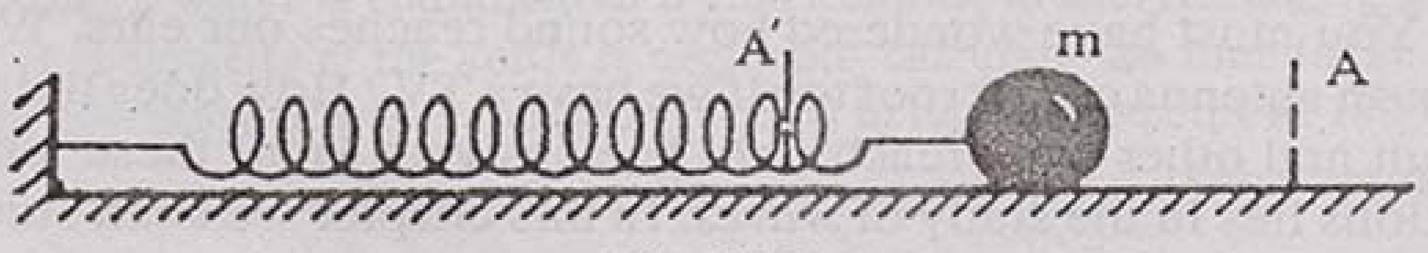


Fig. 12.1

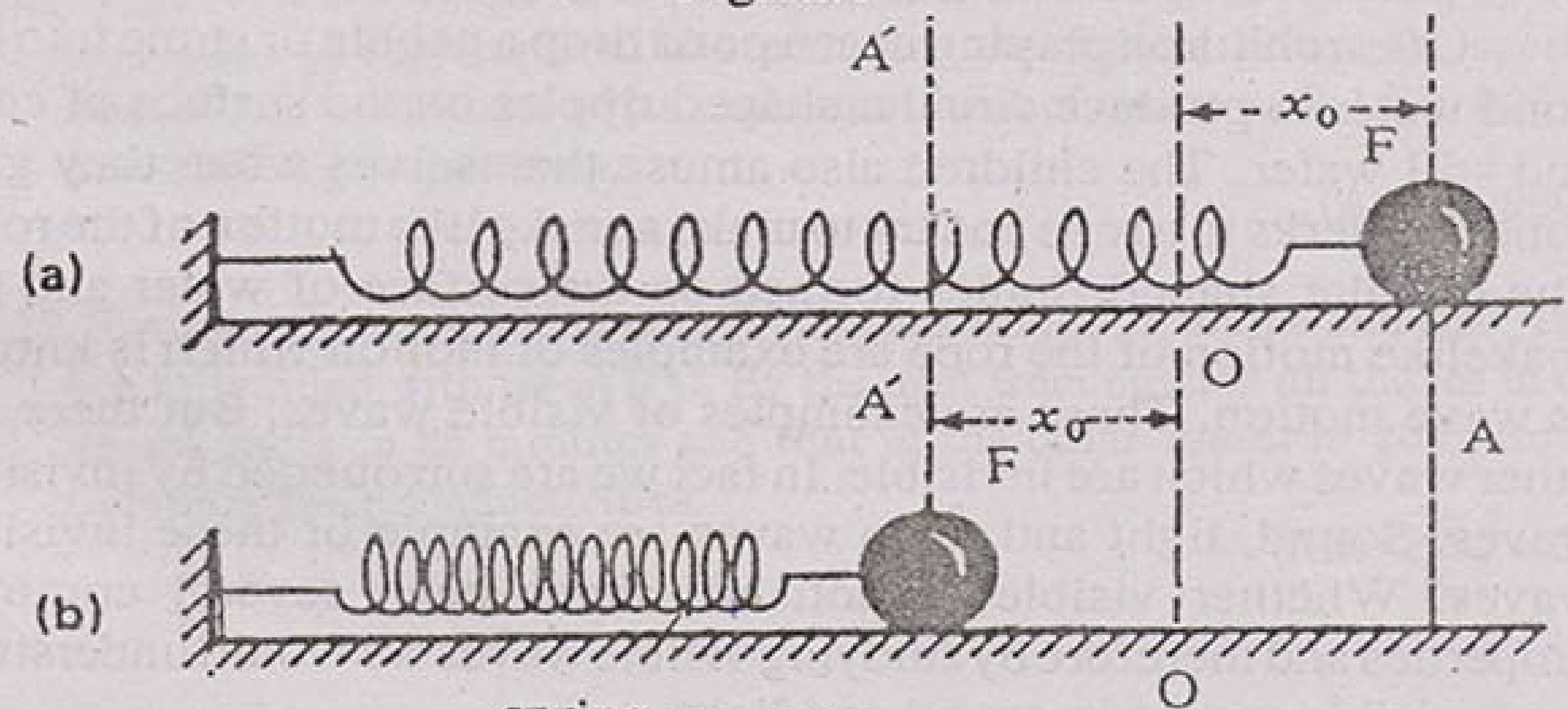


Fig. 12.2

The mass continues this back and forth oscillatory motion until frictional effects cause the motion to stop. All oscillatory motion of this type depend directly on the elastic properties of the vibratory materials.

12.3 SIMPLE HARMONIC MOTION

The vibratory motion of a mass attached to a spring is characteristic of a large and important class of oscillatory phenomenon called simple harmonic motion (SHM), Fig. 12.3 (a) shows a block at rest in its equilibrium position on a frictionless surface. If we apply an external force to displace the block to the right, there will be a restoring force F exerted on the block by the spring and this force is directed to the left, as shown in Fig. 12.3 (b).

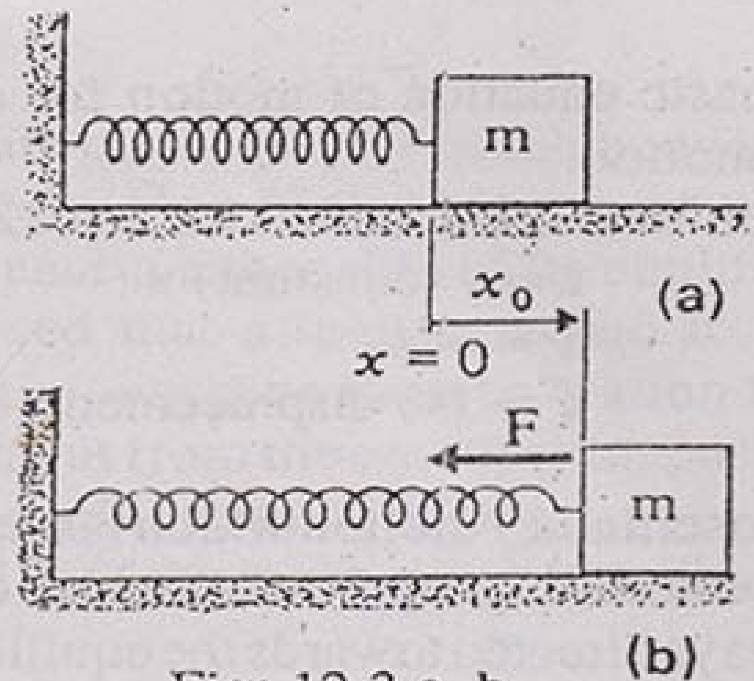


Fig: 12.3 a, b

(a) A mass m rests on a frictionless surface and is attached to a spring. The mass and spring combination is in its normal (equilibrium) condition (b) if m is displaced to the right an amount x_o (by an external force), there will be a restoring force to the left given by $F = \ kx_o$. where k is the force constant characteristic of the particular spring.

We assume that the maximum displacement of the block by the amount x_o has not carried the spring beyond the range of validity of Hooke's law. Then, we can write the relationship between the force and the displacement in the standard form of Hooke's law. But in the discussion of S.H.M. we are usually interested in the restoring force, not in the applied forces. Therefore, to emphasize that the direction of displacement.x, is opposite to that of the restoring force, F we write Hooke's law with a negative sign

$$F = -kx \qquad \text{------------} \qquad (12.1)$$

Where k is proportionality constant usually referred to as spring constant.

When the extension is to the left of the equilibrium position, the restoring force acts toward the right; that is, F has a positive value, whereas x is negative, and eq. 12.1 is still valid.

We know from Newton's equation of motion that the net force on an object must equal the product of its mass and its acceleration. Therefore, at any extension x.

$$F = ma = -kx$$

So that $$a = -\frac{k}{m}x \qquad \text{-------------} \qquad (12.2)$$

This is the basic equation of motion for an object undergoing simple harmonic motion.

Since k and m both are constant the eq. 12.2 can be written as

$$a = -\text{(constant)}\ x$$

Therefore $\quad a \propto -x$

or $\quad a \propto (-)\ \text{displacement}$

This type of oscillatory motion which is characterized by the fact that it has an acceleration proportional to its displacement and the acceleration is always directed towards the equilibrium position (which is indicated by the negative sign) is called simple harmonic motion. The oscillatory particle is then called a simple harmonic oscillator.

We shall now define some important terms used in connection with simple harmonic motion.

(i) Vibration

A vibration means one complete round trip of the body. In Fig. 12.2. It is the motion of mass 'm' from A back to A or from A back to A′ or from O to A to A′ and back to O.

(ii) Time Period (T)

It is the time required to complete one vibration or oscillation. The time period is measured in second

(iii) Frequency (f)

It is the number of vibrations executed by a oscillating body in one second and in S.I unit it is expressed as vibration/s, cycle/s or hertz (Hz).

The frequency f and time period T are related by the equation.

$$f = \frac{1}{T} \qquad \text{-----------} \qquad (12.3)$$

i.e. The frequency is given by the reciprocal of the period of oscillation

(iv) Displacement

Displacement of a vibrating body at any instant is its distance from the equilibrium position at that instant.

(v) Amplitude

It is the maximum displacement of the body on either side of its equilibrium position. In Fig. 12.2 x_o is the amplitude as it is the maximum displacement on either side of its equilibrium position.

We have observed that a mass attached to a spring executes simple harmonic motion because its acceleration is directly proportional to the displacement from the equilibrium position and is always directed towards the equilibrium positions.

The time period T of the mass attached to the spring is given by

$$T = 2\pi\sqrt{\frac{m}{k}} \quad \text{------------} \quad (12.4)$$

and frequency $$f = \frac{1}{T} = \frac{1}{2\pi}\sqrt{\frac{k}{m}} \quad \text{------------} \quad (12.5)$$

Where m is the mass attached to the spring and k represents the spring constant and is equal to the force required to produce an extension of one metre and therefore expressed in N/m.

12.4 EXAMPLE OF SIMPLE HARMONIC MOTION

Simple Pendulum

An ideal simple pendulum consists of a point mass suspended by a weightless and inextensible string from a fixed support. In practice a small metallic bob suspended by a fine string serves the purpose as shown in Fig.12.4

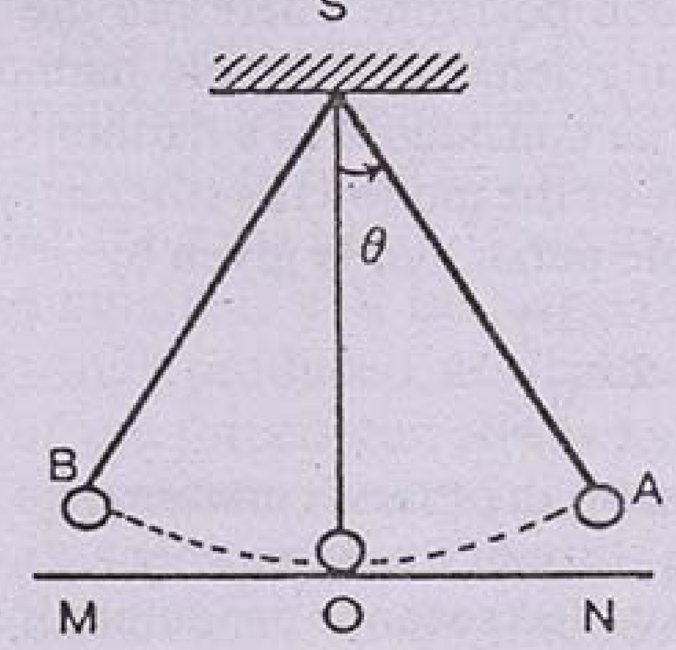


Fig 12.4

If we displace the bob from its mean position O to a new position A and allow it to go it will move towards O under the action of gravity.

The bob will not come to rest at O. but due to inertia it will continue to move towards a point B, while moving from point O to B the bob is moving against the gravity. Its velocity continues to decrease and becomes zero at point B. The bob will then fall back from B to O under the action of gravity. Its velocity continues to increase till it becomes maximum on reaching the mean position O. The bob will not stop at O but continue to move towards A due to inertia. As the bob moves against gravity from O to A, its velocity decreases and becomes zero at A. The bob will then move from A to O under the action of gravity. The whole process is repeated again and again and the bob continues to vibrate between two extreme positions A. and B. The motion of the bob from one extreme position A to other extreme position B and back to position A is called a vibration. The time taken to complete one vibration is known as time period. The maximum displacement of the bob from its mean position, eg; OA or OB is called its amplitude.

As the bob is at its lowest position at point O. its potential energy is minimum here while its kinetic energy is maximum. For that reason its velocity is maximum at the mean position. The bob is at its highest point at either of its extreme positions. Here its potential energy is maximum while the kinetic energy is minimum. Therefore the velocity of the bob at A or B is zero. At all other points between the mean position and the extreme positions the energy of the bob is partly kinetic and partly potential but the total energy at any point during the course of oscillation is constant.

From the above discussion we have observed that the acceleration in simple pendulum is directly proportional to its displacement and is directed towards mean position. Therefore the motion of a simple pendulum can be considered as a simple harmonic motion and the pendulum itself can be considered as a harmonic osscillator.

If the amplitude of the motion is sufficiently small then the time period T of the simple pendulum is given by

$$T = 2\pi\sqrt{\frac{l}{g}}$$

Where l is the length of the pendulum and g is the acceleration due to gravity at the place of the experiment.

A pendulum is called seconds pendulum if it takes two seconds to complete one oscillation. It is used in clocks to record time.

12.5 RESONANCE

Consider a long string or a metallic wire, stretched tightly between two pegs. Four pendulums A,B,C,D of different lengths are fastended to the string, or wire. Another pendulum E of the same length as that of B is also fastened as shown in Fig: 12.5

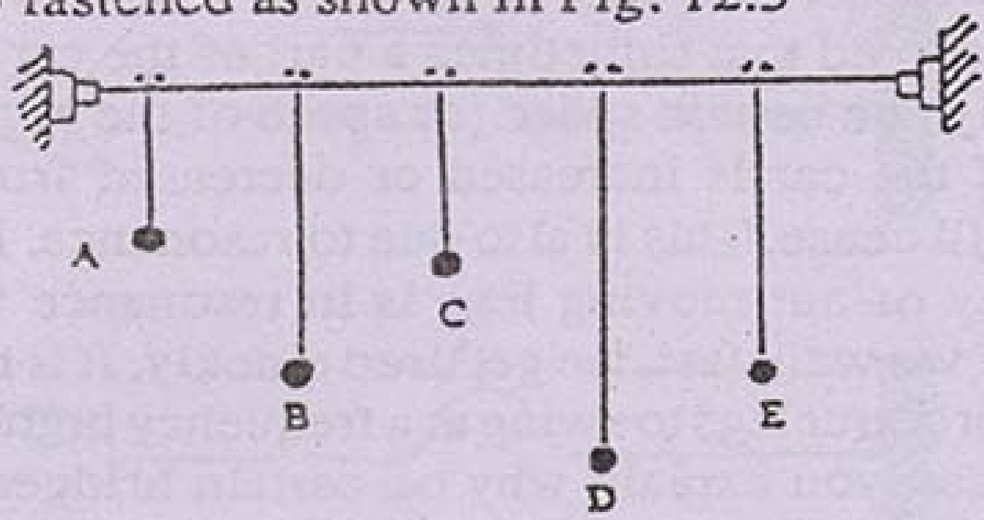


Fig 12.5 Resonance.

When pendulum E is set swinging it will be observed that all the pendulums start to swing but pendulum B begins to vibrate with an increasingly larger amplitude.

As pendulum E is set into vibration it imparts its motion to the string or metallic wire. This string in turn imparts the same periodic motion to the pendulums tied to it. The natural frequency of all other pendulums except pendulum B being different, they do not respond to the same extent to the motion imparted from the string. Pendulum B responds as its natural frequency agrees with the frequency of the motion of the string which in turn was supplied by the vibrating pendulum E. This phenomenon under which pendulum B begins to vibrate is called resonance.

In order to understand resonance let us consider the example of a pendulum hanging freely from a support. If we slightly disturb it, it will begin to oscillate. The time period of these oscillations, called natural oscillations, depends only upon the length of the pendulum. However, if we hold the bob in our hand and move it to and fro then the time period will depend upon the hand. If we move the hand rapidly the time period will be small. It will be large if we move the hand slowly. The vibrations thus produced are called forced vibrations. Now if the frequency of the forced vibrations synchronizes with the natural frequency of the pendulum then the amplitude of the pendulum increases. It is due to this reason that a feeble puff of air blown at regular intervals of time increases the amplitude of the pendulum to a large extent. Thus, "resonance is the response of an object to a periodic

force acting on it. It is greatest when this force has the same period as the object's natural period". Under the influence of this force has the first object not only begins to vibrate but also the amplitude of its vibration increases.

It is observed that sometimes a part of the car begins to vibrate very violently at a certain speed (or speed of the engine) of the car. If the speed of the car is increased or decreased from that value the vibrations will cease. This is also due to resonance. It is observed that the frequency of our moving legs is in resonance with their natural frequency. If we walk fast, we get tired quickly. If is because in the fast motion we force our legs to swing at a frequency higher than the normal frequency. Can you explain why on certain bridges it is written that soldiers must march out of step while crossing the bridge?

12.6 WAVE MOTION

We see that if we dip a pencil into a tub of water and take it out a pronounced circular ripple is set up on the water surface and travels towards the edges of the tub. However, if we dip the pencil and take it out many times, a number of ripples will be formed one after the other. These are shown below in Fig. 12.6.

Fig: 12.6 Ripples on the surface of the water.

If you place a small object a piece of wood, paper etc on the water surface it moves up and down when a wave passes across its position, it does not move outward as the wave do. This shows that the disturbance proceed in water travels outward from the centre of disturbance, the water itself does not move outward. Such up and down movement are vibrations of water which constitute waves are the examples of wave motion.

Waves can also be produced on very long ropes or strings. If one end of the rope is fixed and the other end is given a sudden up and down jerk a pulse shaped wave is formed which travels along the rope see Fig. 12.7.

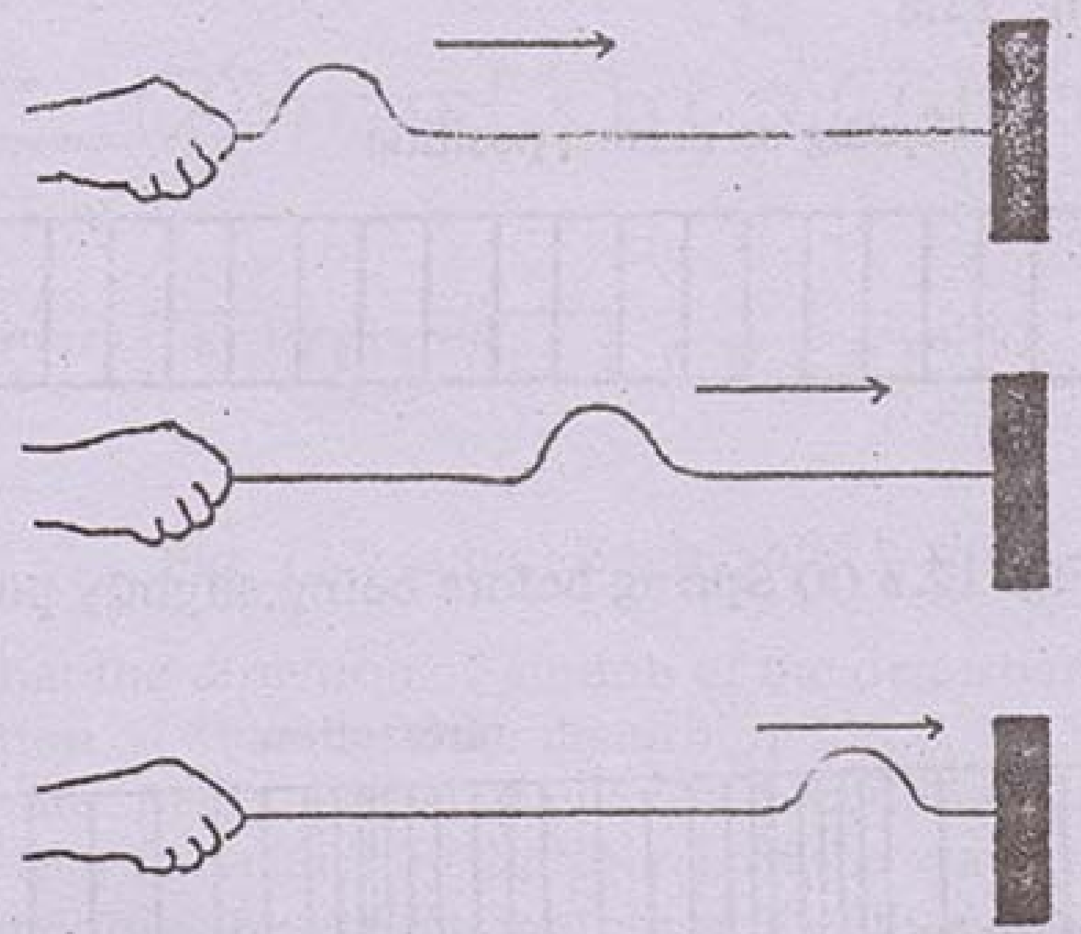


Fig. 12.7 A pulse shaped wave moving along a rope.

If, however, jerking of the rope is continued, a number of wave pulses will be produced. This is called a wave train and it travels along the rope. From the above observations it may be concluded that *A wave is a travelling disturbance.*

Consider a long lossely wound spring, fastened one end against a wall and placed on a smooth frictionless and horizontal table. The windings of the spring are seen to be equally spaced. On slightly pulling the free end of the spring and releasing it we will see that the distance between the winding at the free-end has increased due to the pull of the hand. When the end is released the windings will tend to come back to their normal position due to an elastic restoring force. It is the force which is always acting in such a direction as to pull or push the system back to the equilibrium. At the same time, they drag some of the adjoining windings. Due to this, the windings near the end of the spring come closer together while the distance between the adjacent windings increases. These elongated windings drag the adjancent windings called *rarefactions* or elongations. The regions where the inter-winding distance is less than the normal

called *compressions or condensations.* It will be observed that the compressions and the rarefactions travel along the spring in the direction in which the disturbing force is applied i.e., the pull of the hand. The spring and the compressions and rarefactions are shown below in Fig. 12.8.

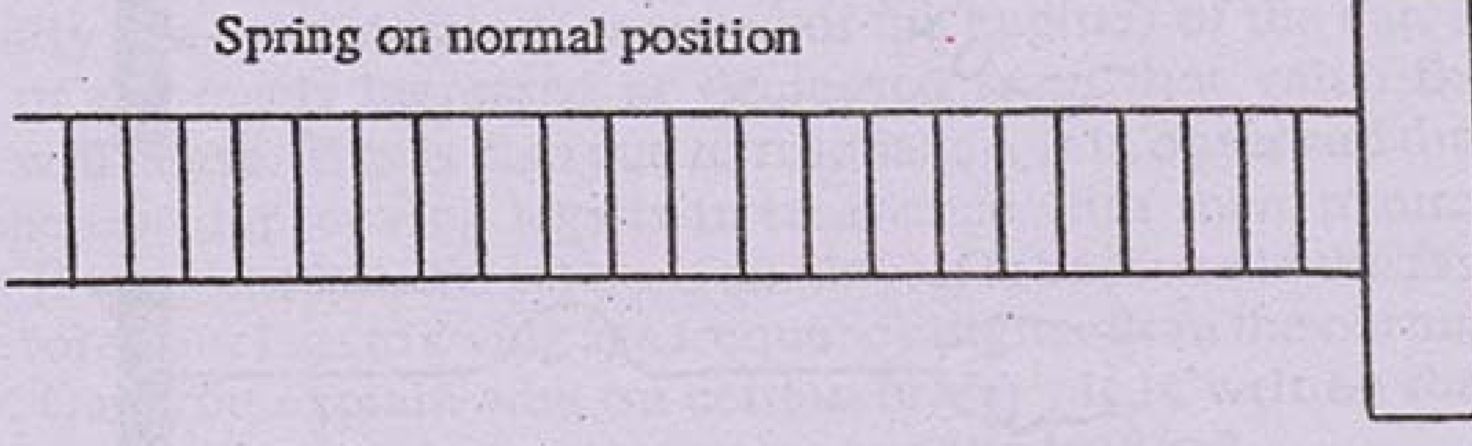


Fig: 12.8 (a) Spring before being slightly pulled.

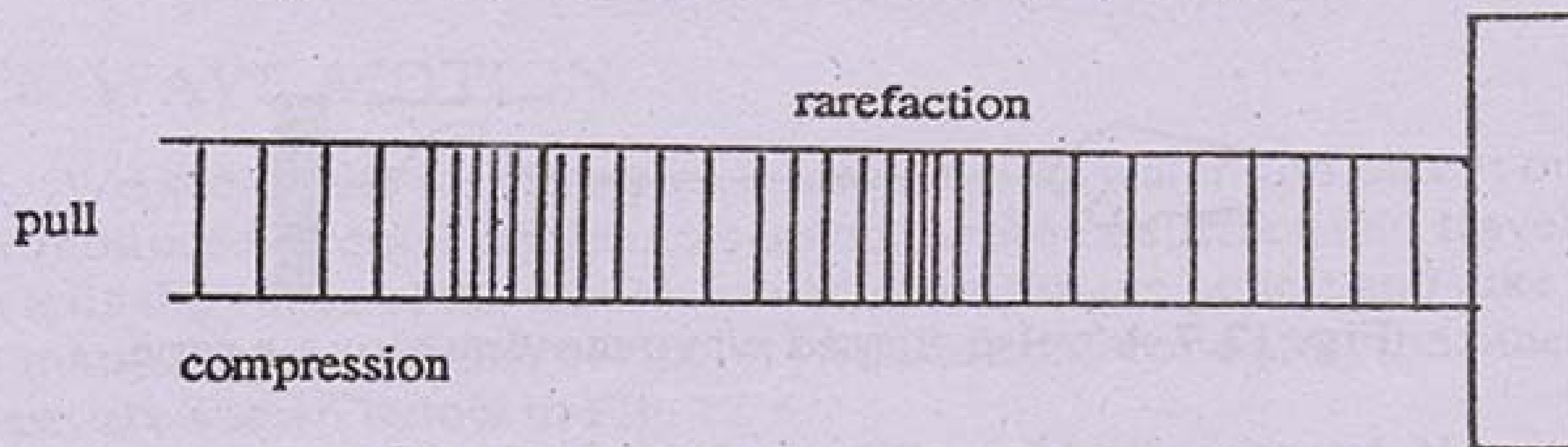


Fig: 12.8 (b) Spring after having been pulled.

The wide spaces indicate rarefactions and close spaces indicate compressions.

Let us now consider another example of a railway engine connected to three bogeys with buffer springs in between them. The engine moves forward a little distance and then stops. Let us examine what happens to the springs in between the bogeys as shown in Fig. 12.9. When the engine moves it elongates the spring between engine E and bogey I. This elongated spring exerts force on bogey I and moves it towards the engine. The spring between engine E and bogey I gets compressed while due to the forward motion of bogey I, the spring between bogey I and bogey 2 gets elongated. It pulls bogey 2 forward to produce a compression between bogey I and bogey 2 and an elongation (rarefaction) between bogey 2 and bogey 3 as shown in Fig. 12.9. c. Here again we observe that rarefactions and compressions (or waves) travel in the same direction in which the disturbing force (the engine) acts.

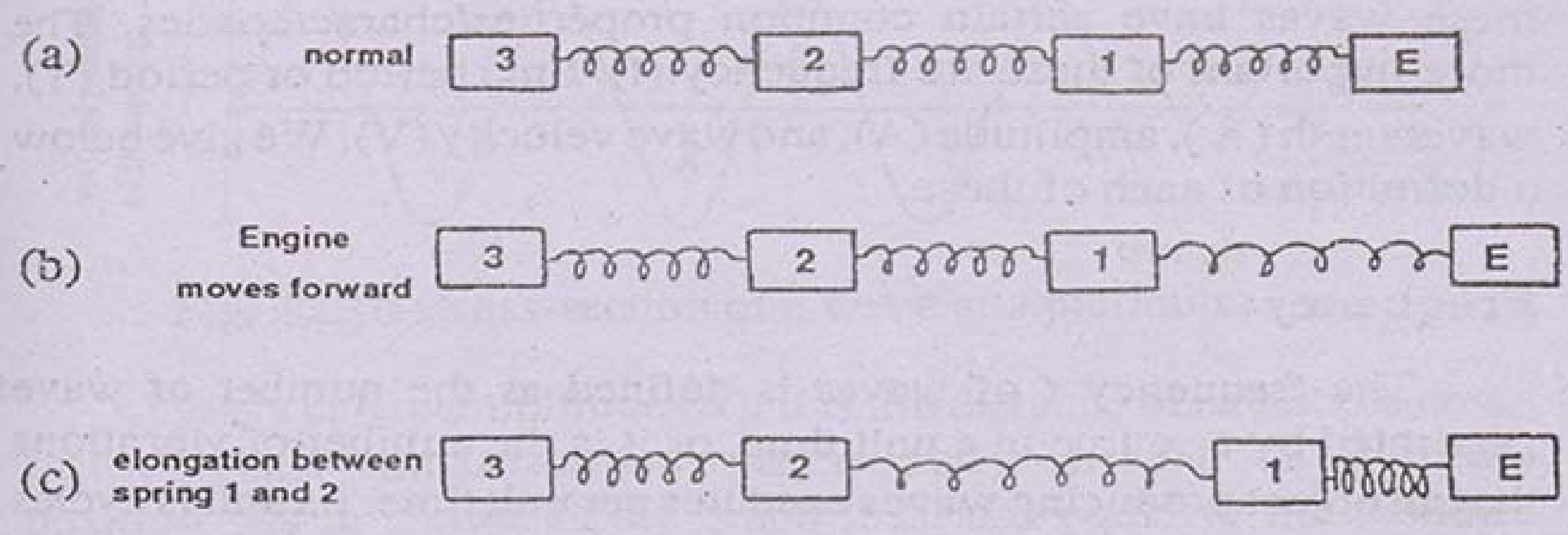


Fig. 12.9 (a,b,c) Engine and bogeys

Note that the direction of motion of the disturbance is the same as the direction of motion of the disturbing force.

There are many kinds of waves depending on the kind of disturbance that is created by the source of the wave and the kind of material or medium the wave trávels through. It is useful, therefore, to introduce two basic types of waves; *transverse and longitudinal waves*. The waves formed on the water surface and the waves formed on the rope are example of transverse waves, while the waves formed when a spring is pulled harizontally and those demonstrated by engine and bogey are examples of longitudinal waves.

Thus a "transverse wave is one in which the particles of the medium or the disturbance travel in a direction perpendicular to the direction in which the wave travels". Radio waves, light waves and mocrowaves are transverse waves.

The second category of wave which is called a longitudinal wave can be seen in the spring experiment and the engine and bogey experiment. In this experiment the displacement of the particles of mediumis parallel to the direction of the wave motion. Thus a longitudinal wave is one in which the disturbance is parallel to the line of travel of the wave. *A sound wave is a longitudinal wave.*

12.7 CHARACTERISTICS OF A WAVE.

The transverse and longitudinal waves are periodic waves since these consists of patterns that are produced over and over again by the simple harmonic motion of the agent creating the disturbance. Therefore

these waves have certain common properties/characteristics. The more important of these are frequency (f), time period or period (T), wavelength (λ), amplitude (A), and wave velocity (V). We give below a definition of each of these.

Frequency

The frequency f of waves is defined as the number of waves generated by a source in a unit time, or it is the number of vibrations that an object producing waves executes per unit time. Its unit is cycles per second (cs^{-1} or c/s) or hertz (Hz). It is given as $f = \frac{1}{T}$.

Time Period

The symbol T (Fig 12.11) on the graph represents the time period or simply the period T. It is defined as the time taken by an oscillating particle to complete one up and down cycle as one cycle of the wave (from the top of one crest to the top of the following crest) passes by it. This is equivalent to saying that the period is the time required for the wave to travel a distance of one wavelength.

As for any simple harmonic motion the period T is related to the frequency f, according to the relation.

$$\text{Time period} = \frac{1}{\text{frequency}}$$

$$= \frac{1}{f}$$

The wavelength and Amplitude

In order to make this terminology clear, we make use of the following graph of a transverse wave (Fig. 12.10). The straight line on the abscissa shows the normal state of the spring. The symbol A on this graph represents the maximum displacement of a particle of the medium from its undisturbed position.

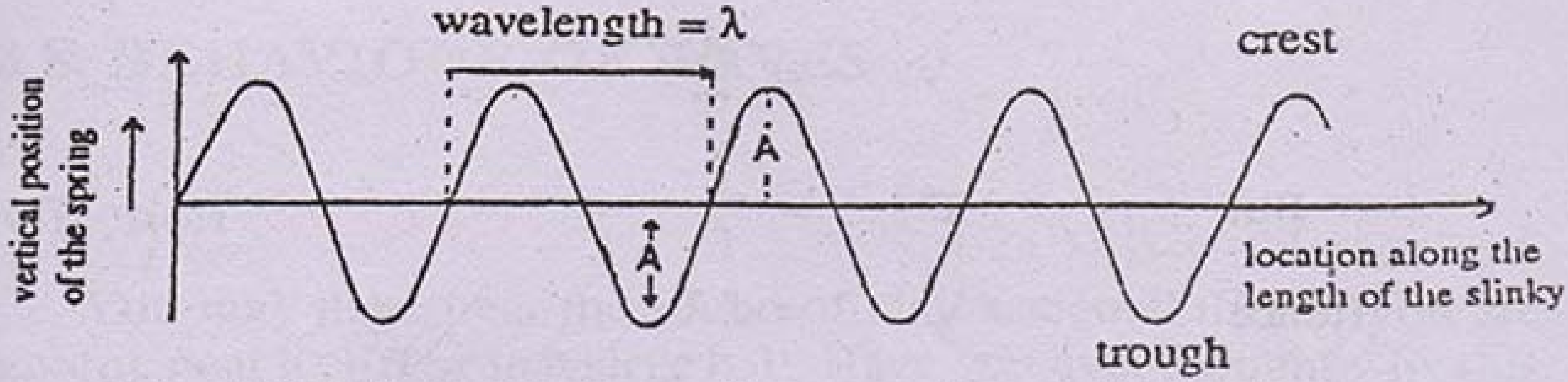


Fig. 12.10 Cross-section of a wave at a particular time.

This is called amplitude (A). It is the distance between the crest, i.e the highest point on the wave pattern, and the normal (undisturbed) position. As in shown in the graph it is also equal to the distance between the *trough* i.e the lowest point on the wave pattern, and the normal position.

The wavelength λ is the horizontal distance covered by the wave in one complete vibration of the particle or it is the distance between two consecutive crests or between two consecutive troughs, or any two successive equivalent points on the wave. It is also the distance between two consecutive compressions or between two consecutive rarefactions (elongations) of a longitudinal wave.

Now if we just confine our attention to the up and down movement of a particular particle as the wave (disturbance) passes and consider it as something that changes with time we get the graph of disturbance or displacement (y) plotted against time (Fig.12.11)

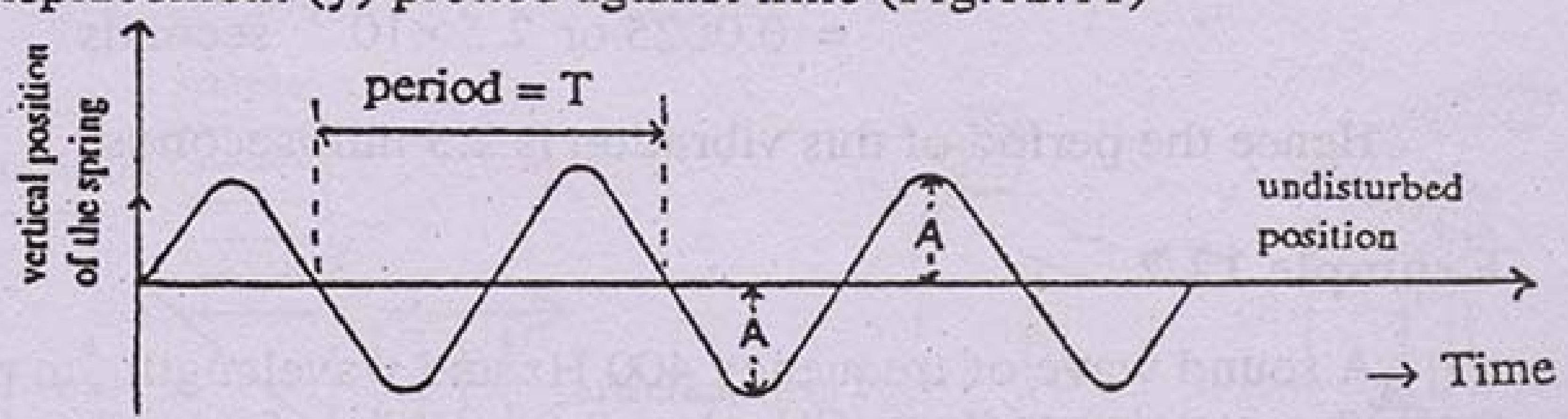


Fig. 12.11 Displacement at a particular location

Wave Velocity

A simple relation exists between the wavelength the period and the velocity of a wave. The speed of a wave is the distance travelled by it in a unit time (in the direction of wave).

$$\text{Velocity} = \frac{\text{Distance}}{\text{time}} = \frac{\text{wavelength}}{\text{time period}}$$

$$V = \frac{\lambda}{T}$$

But $$T = \frac{1}{f}$$

Hence $$V = f\lambda$$

Waves of the same type whatever their wavelength of frequency, travel with the same velocity in the same medium. For example, light and radio wave travel through air at about 3.0×10^8 m/s. Sound waves travel through air at about 330 m/s at O°C.

Example 12.1

A sitar string vibrates at 400 Hz. What is the time period of this vibration?

Solution

Frequency, $f = 400$ Hz

To find the time period T we use equation,

$$T = 1/f$$
$$= 1/400s^{-1}$$
$$= 0.0025 \text{ or } 2.5\times10^{-3} \text{ seconds}$$

Hence the period of this vibration is 2.5 milliseconds.

Example 12.2

A sound wave of frequency 400 Hz and wavelength 3m passes through a certain medium. Calculate the velocity of the wave in that medium.

Solution

The frequency $f = 400$ Hz

and the wavelength $\lambda = 3m$

Using the equation $V = f\lambda$

We get the wave velocity $V = 400s^{-1} \times 3m$

$= 1200\ ms^{-1}$

12.8 BEHAVIOUR OF WAVES

Reflection

You may have hear the sound of clap a second time if you are standing near a cliff or in a large hall. Have you ever thought why this happen? You might have observed that water waves turned back. Similarly the second sound of the clap is due to the bouncing back of the sound from the surface of the cliff or the distant wall in a hall. The interval between clapping and hearing the clapping sound depends upon the distance between the person clapping and the reflecting surface.

Similarly children play with mirrors flashing sunlight into the eyes of the other children fig 12.12. Can you tell how these children direct sunlight into the eyes of even those who are hidden from the sun?

Both the activities described above, show that a wave can be bounced back from a surface. This bouncing back of a wave from a surface is called reflection. The angle at which the wave is reflected is equal to the angle at which the wave is incident on the surface. Waves coming from the source and hitting an obstacle or barrier are called incident waves. Those that seem to originate from the barriers, called reflected waves, have the same frequency because they are produced by the same source in the same medium having uniform depth.

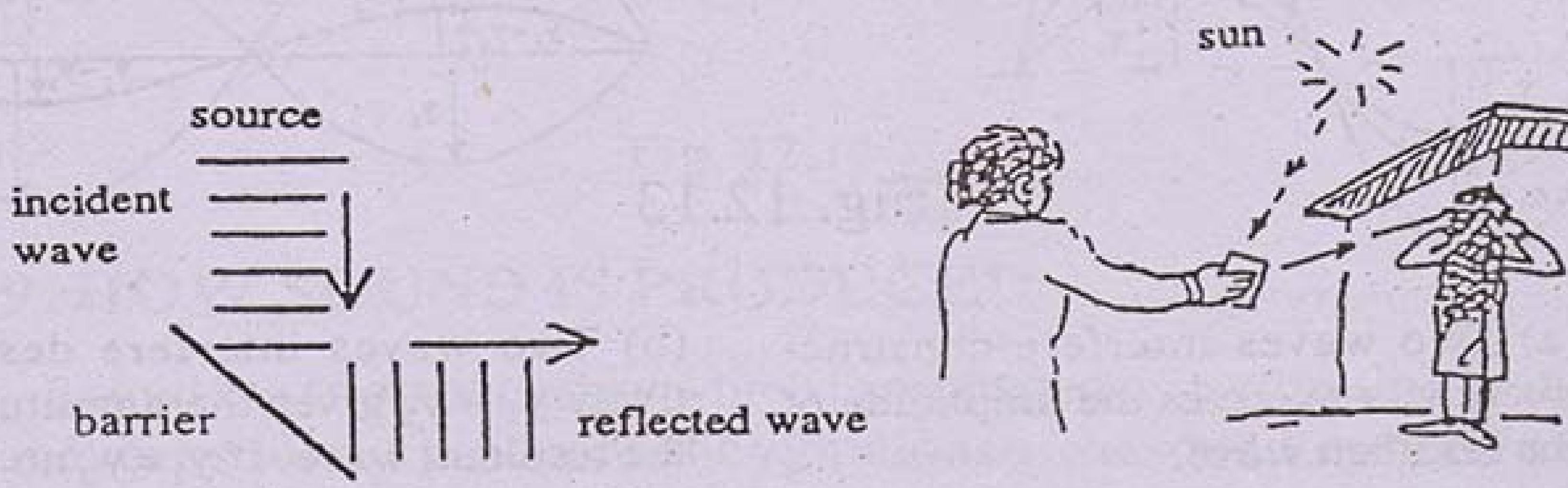


Fig.12.12 Reflection of a wave

In the second example, the light is reflected by the mirror. By tilting the mirror in the appropriate direction the sun's rays can be thrown in many directions. The longitudinal waves such as sound waves obey the same laws of reflection as are obeyed by transverse waves such as light waves.

Interference

What happens when two sets of waves formed by dropping two stones in a pond on meeting?

Suppose two different sets of waves of the same amplitude formed on the surface approach each other. The waves may have started from the opposite ends of a string or they may be two circular waves formed on the surface of water when two stones are dropped in it. When two sets of waves meet, they are neither reflected nor absorbed by each other; one simply passes over the other. However, at those points where the waves meet, the net amplitude of the combined wave will be the algebraic sum of the displacements of the two separate waves. By interference we mean the interaction of two waves passing through the same region of space at the same time. It may be observed that if at a given point the crests or the troughs of the two waves arrive simultaneously then the combined wave is larger than either of the two waves. This is called constructive interference (Fig. 12.13 a). If, however, the crest of one wave arrives simultaneously with the trough of the other wave then the two will cancel each other and no wave will be observed. This is called destructive interference (Fig. 12.13 b).

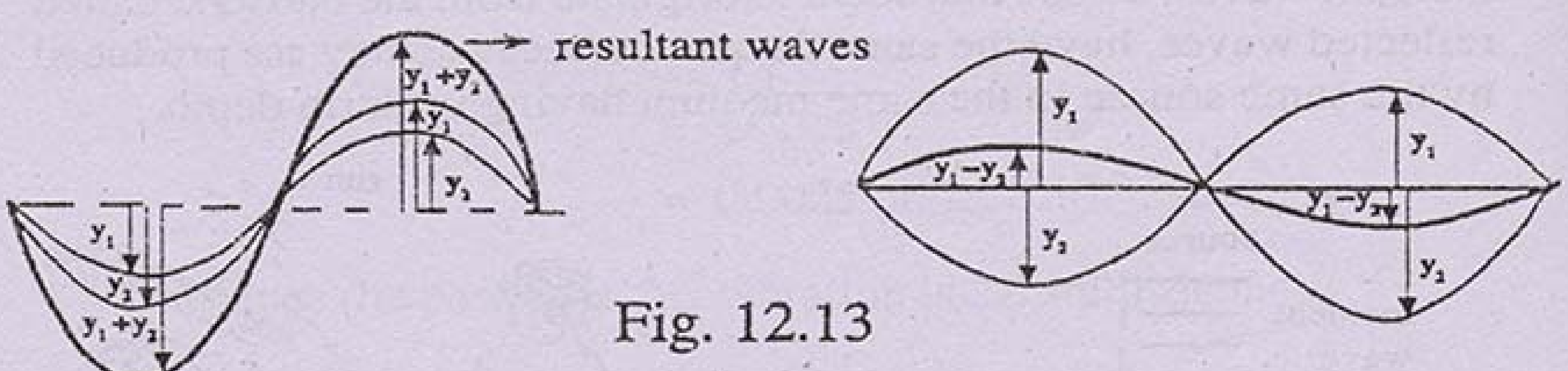


Fig. 12.13

(a) Two waves interfere constructively $y_1 + y_2$ gives the amplitude of the resultant wave.

(b) Two waves interfere destructively $y_1 - y_2$ gives the amplitude of the resultant wave. If $y_1 = y_2$ no wave will be observed.

Stationary Waves

In the interference phenomenon, as discussed in the previous article, the amplitude and the frequency of the interfering waves are often different. If, however, two waves of the same amplitude and frequency travelling in opposite direction meet one another, the resulting interference pattern gives rise to what are called standing waves or stationary waves. It is easy to analyze and study waves with the help of standing waves.

Stationary waves may be produced by means of a single source. For example, the incident waves may interfere with the reflected waves to produce stationary waves. As the incident waves and the reflected waves originate from the same source, pass through the same medium in they possess the same amplitude and frequency. Such a wave is called a "stationary wave" because it does not appear to be moving. The points of destructive interference, called nodes, and of constructive interference called antinodes; remain in fixed positions. Stationary waves are formed at more than one frequency. These are shown below in Fig.12.14. The distance between a node and an antinode is one quarter of the wavelength.

Thus we can say that:

A standing (stationary) wave is produced when two waves of the same amplitude and frequency, that are travelling in the opposite directions, are combined.

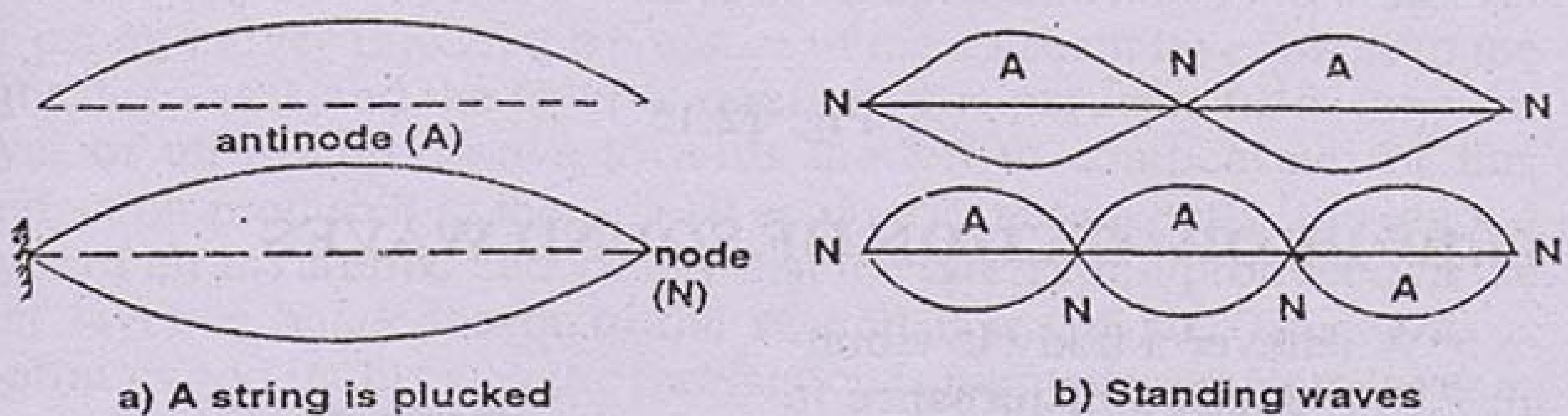


Fig. 12.14

12.9 HOW SOUND IS PRODUCED

Sound is a form of energy which is produced by a vibrating body. Whenever a sound is produced, the vibrations of the sounding body can be seen or felt. For example, a metallic vessal is struck with a spoon, it starts giving out sound and the vibrations of the vessal can be felt by touching it gently with hand. As another example, strike a tuning fork on a rubber pad. A peculiar sound is produced and the prongs of the tuning fork begin to vibrate. These vibrations can be demonstrated by bringing a pith ball suspended by a thread near one of its prongs. The pith ball will fly away as soon as it touches the prong (Fig. 12.15). If we switch on a radio and make its sound louder, its cabinet will start vibrating. If we place small pieces of paper on its cabinet, they will start

vibrating. From these observations we can conclude the sound is produced only if a body is vibrating.

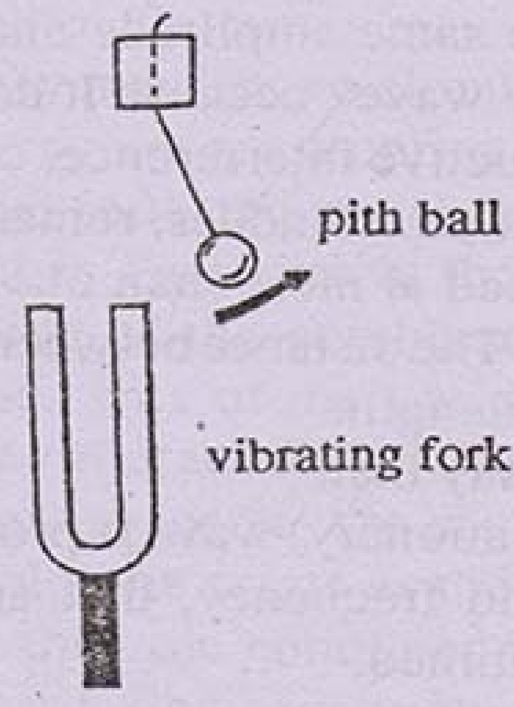


Fig. 12.15

12.10 PROPAGATION OF SOUND WAVES

Whenever a body is vibrating, it produces a disturbance in the surrounding air. This disturbance reaches our ear in the form of waves thus producing the sensation of sound. If their is no medium around the vibrating body, the sound waves will not reach our ear and we will not hear any sound. Let us perform an experiment to demonstrate the above mentioned fact.

Suspend an electric bell in a jar by its wires through a cork fixed in its mouth as shown in Fig. (12.16). Switch on the bell. We will hear the sound of the bell. Now start removing air from the jar with the help of an exhaust

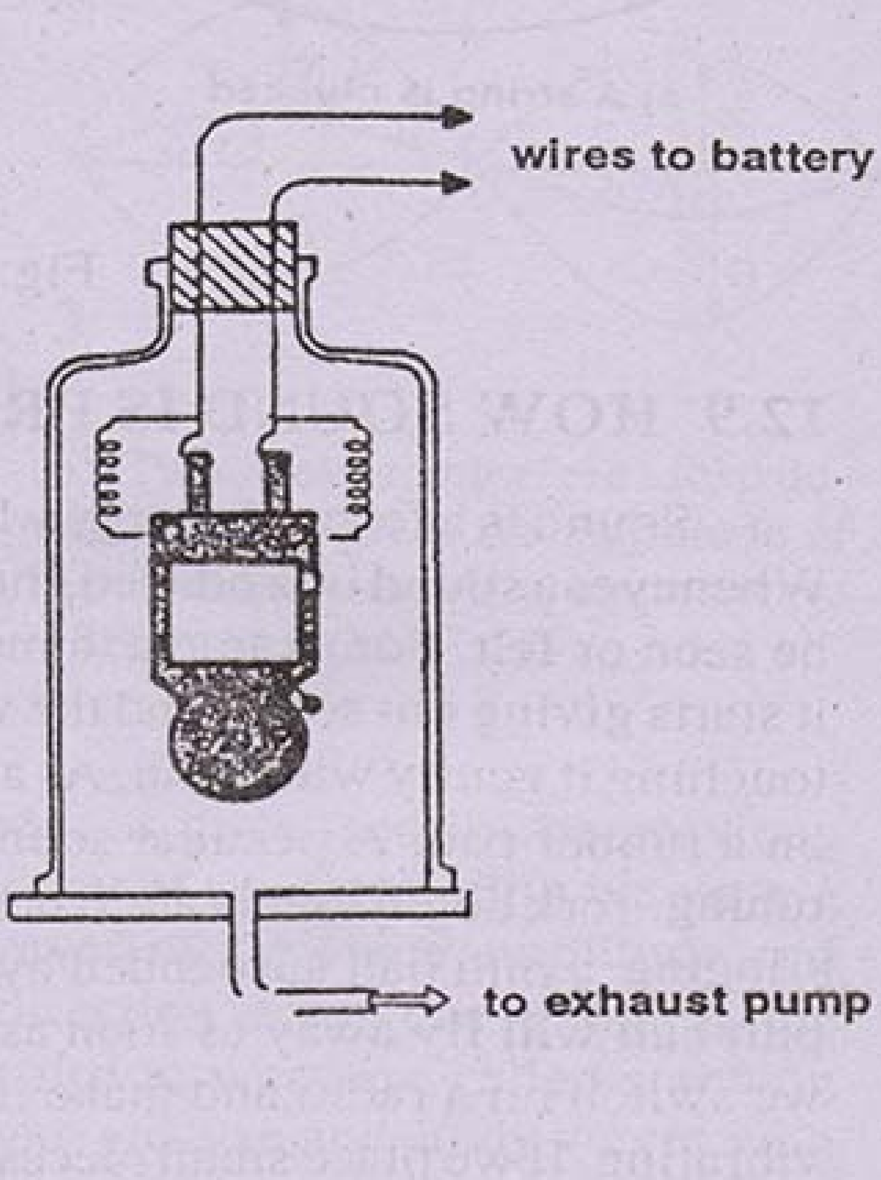


Fig. 12.16

pump. The loudness of the sound of the bell will start decreasing ultimately, although the hammer is still seen striking the bell. This experiment shows tht air is necessary for the propagation of sound infact a material medium such as air, water metals etc., is needed for the sound to travel from one place to another.

A vibrating body produces sound and it travels from one place to another in the presence of a medium. How the sound waves travel in a medium? To answer this, let us consider the vibrations of the prongs of a tuning fork.

A sound is produced when we strik a tuning fork with a rubber pad and its prongs start vibrating. In Fig. (12.17), the right prong of the tuning fork is vibrating betweel the positions A and B about its mean position O. When the prong moves from A to B, it compresses the adjacent layer of air to the right. This compressed layer of air hands over its compression to the next this way the compression is handed over from layer to layer and a wave of compression moves forward. As the prong moves B to A, the pressure of the adjacent layer of air to the right decreases and the layer is rarefied. The particles from the next layer of air slightly move towards the left to compensate for this rarefaction produced in the first layer of air. Consequently, the second layer of air is rarefied and we say that the rarefaction produced in the first layer is handed over to the second layer of air. This process continues and in this way the rarefaction is harmed over from layer to layer, and a wave of rarefaction moves forward. Thus as the strip vibrates to and fro, compressions and rarefactions are produced one after the other and they travel outward. The series of compressions and rarefactions is known as sound waves. In Fig. (12.17), the compression is represented by C and rarefaction is represented by r. Notice that molecules of air vibrate about their mean positions as the sound waves pass through the air. We have noted that molecules in air in a layer move forward as compressions passes over it, but they move backward as rarefaction move through the layer. It is quite clear that the molecules of air vibrate in the direction in which the wave moves. Thus the sound waves are longitudinal.

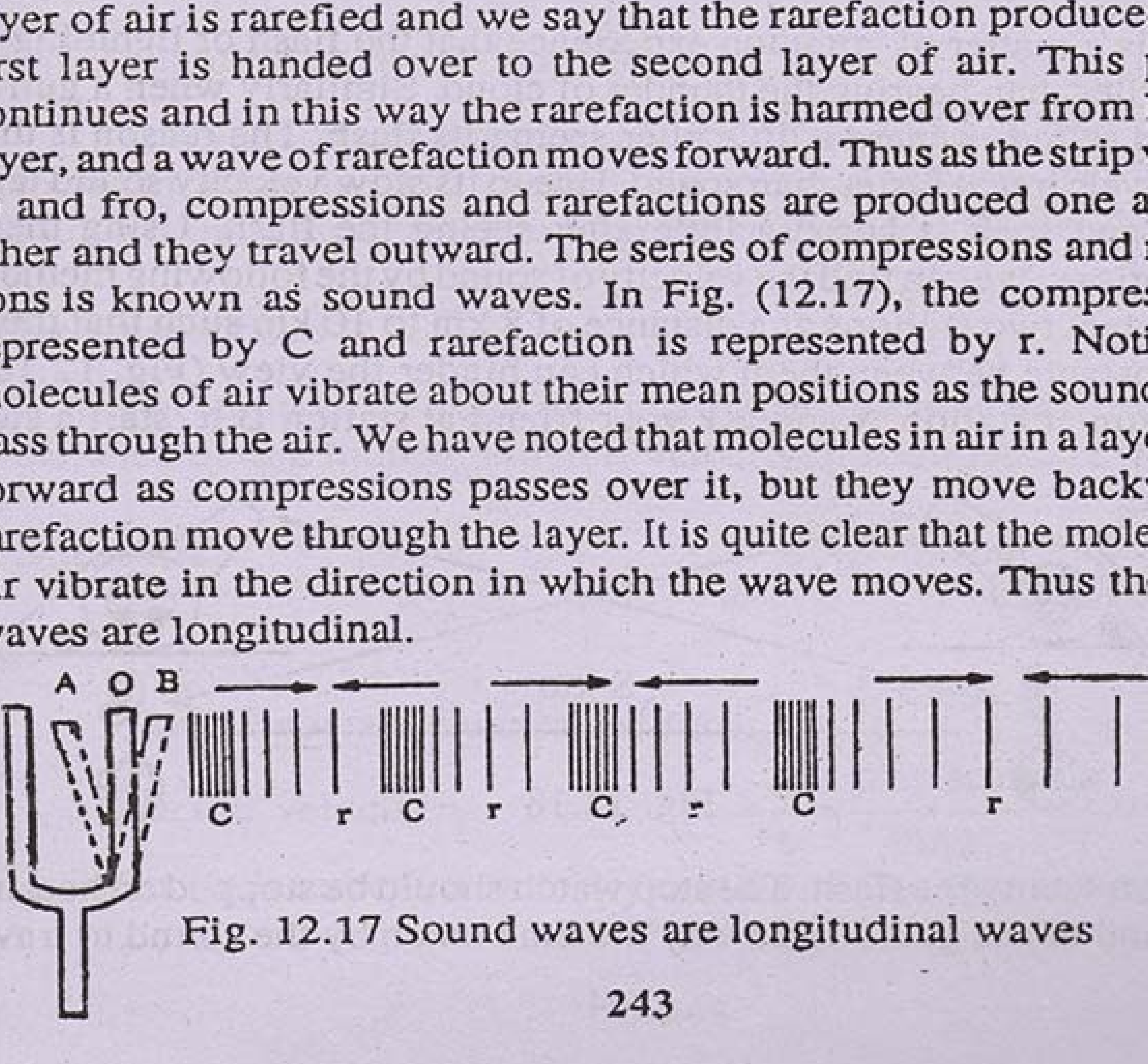


Fig. 12.17 Sound waves are longitudinal waves

When the sound waves strike the ear drum it starts vibrating and, we hear the sound. Thus three things are necessary for the sound (i) vibrating body (ii) some material medium like air, water etc. and (iii) receiver like ear.

12.11 AUDIBLE FREQUENCY RANGE

We know that a body producing sound is always vibrating. But why the sound is not heard when a simple pendulum is vibrating. The reason is that its frequency is very low. Our ear can hear only those sounds whose frequency is between 20 Hz to 20,000 Hz. That is, the ear can neither hear a sound of frequency less than 20 Hz nor a sound of frequency greater than 20,000 Hz. A sound of frequency greater than 20,000 Hz can be produced but human ear cannot detect it because ear drum cannot vibrate with such a high frequency. The second having a frequency more than 20,000 Hz is known as ultrasonics. The audible range (20 Hz to 20,000 Hz) is different for different persons and it also varies with the age.

12.12 VELOCITY OF SOUND

It is a matter of common experience that the flash of lightning is seen earlier than hearing the thunder of cloud. Similarly when a gun is fired, its sound is heard a little after seeing its flash. The reason is that light travels much faster than sound. Due to its slow velocity sound lags behind and so it is heard a little after seeing the flash. Using these observations, we can find the velocity of sound by the following method.

Select two stations at a distance of 8 km to 10 km such that there is no obstacle between them which can hinder the view (Fig. 12.18). Fire a gun at station A and ask your friend at station B to start a stop

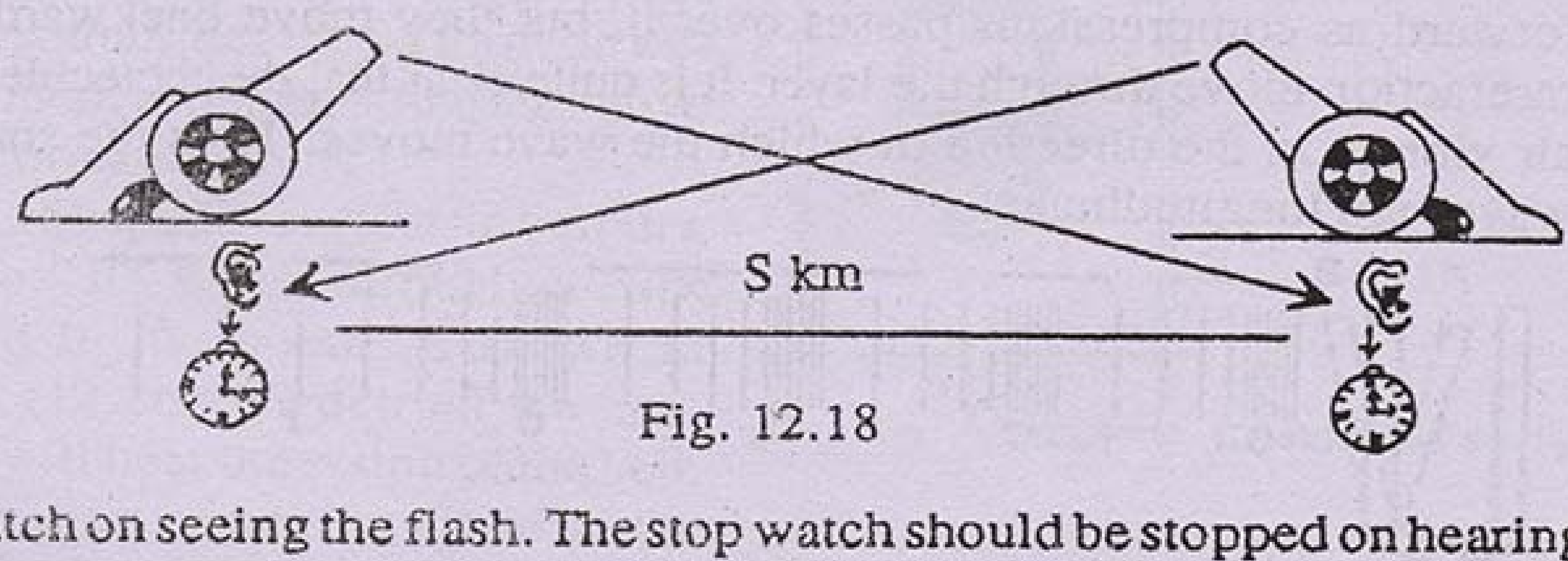


Fig. 12.18

watch on seeing the flash. The stop watch should be stopped on hearing the sound of the gun. In this way the time taken by the sound to travel

from station A to station B is measured. The distance S between the two stations is already known. So the velocity v of the sound can be calculated by the formula:

$$V = \frac{S}{t}$$

The velocity of air is a source of error in this experiment. The apparent velocity of sound will be more than its actual value if the air is blowing from station A to staion B and it will be less than actual velocity if the air is blowing from station B to station A. This source of error can be eliminated by first firing the gun at station A and measuring the time t_1, between seeing the flash and hearing the sound of the gun at station B and then by firing the gun at station B and measuring the time interval t_2 between seeing the flash and hearing the sound at station A. The mean of these time intervals is calculated to find the exact value of velocity of sound.

The distance between the two stations A and B is S

Average time $t = \frac{t_1 + t_2}{2}$

Velocity of sound $V = \frac{S}{t}$

The vellocity of sound in air at O°C is 330 m/s. The velocity of sound in water is 1450 m/s and in iron it is 5130 m/s. The velocity of sound is increased with the increase in temperature.

Example 12.3

The thunder of a cloud is heard 7s after seeing the flash. Find the distance of the thunder cloud if velocity of sound is 341 m/s.

Solution

Velocity of sound = 341 m/s

Time between thunder and flash = 7s

As the velocity $v = \frac{S}{t} = \frac{\text{displacement}}{\text{time}}$

So diplacement = velocity × time = 341 m/s × 7s. = 2387 m

12.13 CHARACTERISTICS OF SOUND

We hear different sounds all around us. Some of these, like the sound from a flute, a piano, and a sitar and the voice of cuckoo are pleasant to hear while the cawing of a crow, the bleating of sheep, the howling of winds, the ratting sound of roadrollers or trucks are sounds which are jarring to hear. The former of these sounds are called musical sounds while the latter are known as noise. In musical sounds there is a regularity in the variation of frequency and amplitude. Noise on the other hand has abrupt changes in amplitude and frequency and there is no regularity in the variation of frequency. For this reason we shall confine our studies to musical sounds. A musical sound has the following characteristics (a) Loudness (b) Pitch and (c) Quality or timbre.

Loudness

Loudness of sound depends upon the intensity of the sound waves. In tensity of sound waves is defined as the energy carried by the sound waves through a unit area placed perpendicular to the direction of propagation of waves per second. Loudness enables us to distinguish between a faint and a loud Sound. It is actually a sensation of human consciousness; However the loudness of sound depends upon the following factors.

(i) *Area of the vibrating object:* A school bell has a large area and therefore it produces a loud sound as compared with the sound of a house bell. The drum (Dhool) produces a loud sound compared with that produced by a dhoolac (a smaller drum). Thus "the larger the area of the vibrating object,the louder will be the sound produced."

(ii) *Amplitude of motion of the vibrating object:* Strike a drum lightly first and then strongly. Listen to the sound produced in both cases. When is sound produced faint and when is it loud? The sound is faint when the drum is struck lightly. In this case the membrane of the drum has a small amplitude through which it vibrates. When the drum is struck strongly the membrane of the drum vibrates through a greater amplitude and hence we hear a loud sound. Thus "the greater the amplitude of vibrating object the louder the sound produced".

The loudness of a sound also depends upon the distance of the sources of sound from the listener. We move away from a drum being beaten violently to avoid the uncomfortably loud sound. We also

experience that a sound is heard to be louder if it travels in the same direction in which the wind blows. It seems to be faint if it travels in the opposite direction.

Pitch

We often hear the whistle of a train, the cry of a baby, the chatter of a child and the talking of a man or woman. Some of these sounds are shrill and some are flat. This property of the sound by virtue of which we can distinguish between a shrill sound and the flat sound is called the pitch of the sound. The pitch of the sound depends upon the frequency of the sound. The greater the frequency, the higher the pitch and vice versa. This can be demonstrated by an activity described below.

Take a disc which is capable of rotation about an axle. Drill holes of equal size and equally spaced along the periphery of the disc as shown in Fig.12.19 Take a long rubber tube connected to an air compressor at one end and to a metallic jet on the other. Pass a rubber belt over the axle of the disc and also over a wheel which can be rotated with a handle.

Rotate the handle to set the disc in motion. Direct the jet of compressed air onto the holes. When a hole comes against the jet the air passes through it and disturbance is produced. As the rotation of the disc increases the number of times the air passes through the holes also increases and hence the frequency of the sound. The pitch of the sound rises. If now we slow down the disc the ptich will become low, the sound will become flat and the frequency will decrease.

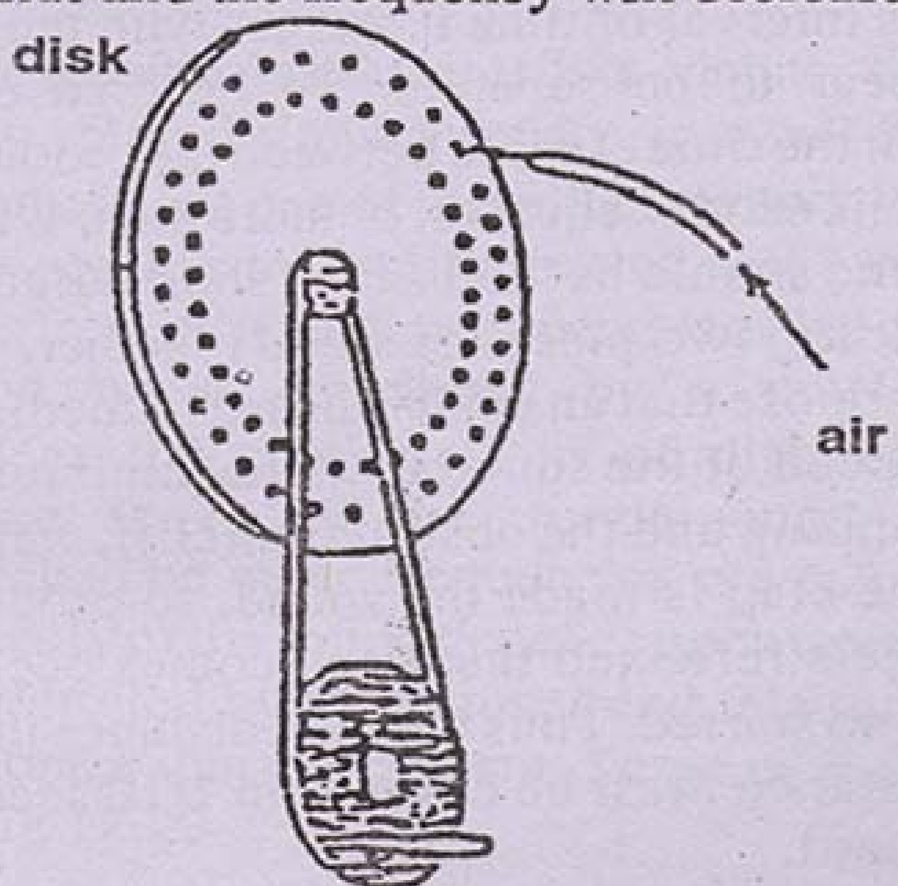


Fig. 12.19 A perforated disc through which air passes.

Example 12.4

There are 48 holes in a disc siren arranged in the form of a ring. The disc rotates at 400 revolutions per half minute. What is the frequency of the sound emitted by an air jet placed against the holes.

Solution

Number of holes in the disc $= 48$

Speed of rotation of disc $= \frac{400}{30}$ revolution per second

Frequency of sound emitted $= \frac{400}{30} \times 48 \text{ Hz} = 640 \text{ Hz}$

You may carry out similar experiment using a piece of cardboard positioned on the frame of the bicycle so that it just touches the spokes of the wheel. Every time a spoke passes the cardboard it strikes it and makes a noise. The pitch can be changed by cycling faster or by moving the cardboard closer or further away from the axle of the wheel.

Echo

The sound heard after reflection from a surface is called an Echo. In a normal human ear the effect or the sensation of sound persists for 1/10th of a second after the sound has ceased. If some sound enters the ear within this interval of time it merges with the previous sound and does not appear to be separate. To hear an echo it is therefore necesssary that the time elapsed between the production of a sound and the hearing of its echo is equal to, or more than, 1/10th of a second. For clarity the sound should be loud and of short duration such as clapping of hands or hitting two pieces of wood together.

Let us suppose that an echo is heard immediately after clapping. Let v be the speed of the sound then the minimum distance between the person clapping and the obstruction(cliff, wall, building etc) is d. Now when the clap is made the sound, so produced, travels to the obstruction, gets reflected there and comes back to the place from where the sound started. Thus the total distance that the sound travels is 2d. This distance must be covered in 1/10 second or more so that the echo is heard.

Hence $\quad 2d = V \times 1/10$

The velocity of sound in air at 15°C is about 340 ms^{-1}

Thus, $2d = 340\ ms^{-1}\ 1/10s$

Hence, $d = 17\ m$

Thus the echo of a sound of short duration, like that of a clap, can be heard distinctly if the minimum distance between the person and the obstacle is 17 m.

Example 12.5

A person wants to find the width of a lake by shouting and listening for the echo from the far side. If the echo returns after 2 seconds how wide is the lake? (Velocity of sound is 340 ms^{-1}

Solution

Speed of sound $= 340\ ms^{-1}$

Time elapsed between shouting and echo $= 2s$

Thus $2d = V \times t$

or $2d = 340\ ms^{-1} \times 2s$

and $d = \dfrac{340 ms^{-1} \times 2s}{2} = 340\ m$

Hence the width of the lake will be = 340 m

Quality or Timbre

If different musical instruments like violins, flutes , sitars and cymbals are being played simultaneously in a hall it is not difficult to identify the sound of each one of these. This property of the sound by virtue of which it is possible to identify a sound of the same loudness and pitch but originating from different instruments is called quality or timbre.

The sound waves produced by a musical instrument may be regarded as the combinations of different frequencies. The simplest, one is called the fundamental frequency. It determines the pitch. The other frequencies are called overtones or harmonics. The quality of the sound depends upon the wave form of the resultant and is controlled by the number and relative intensities and phase of the harmonics that are present. Some of the waveforms are shown below in Fig 12.20

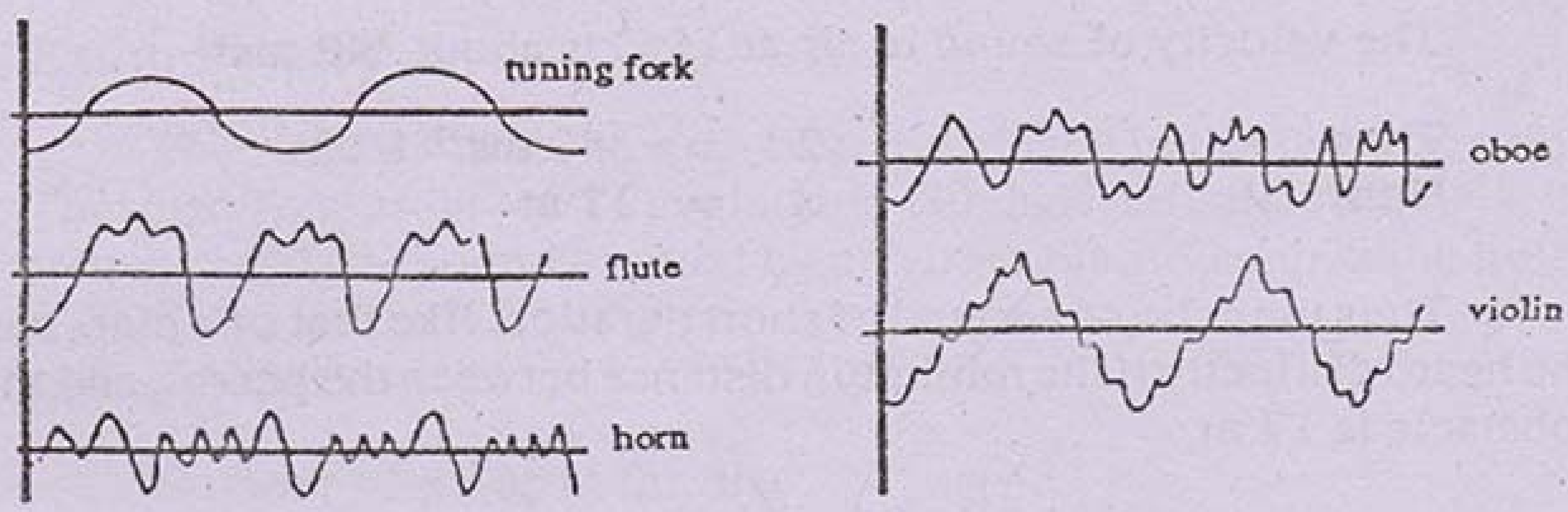


Fig. Waveforms of notes from some musical instruments.

It is the quality of sound that enables us to recognise the sound of an individual person whom we know. Nature has such a great diversity that it is very rare to have the overtones of two persons exactly the same. So this difference in the overtones makes it possible to recognise the sound and hence identify the person.

12.14 REFLECTION OF SOUND

As already described waves are reflected at the surface separating the medium, in which the wave travels, from some other medium at which the wave is incident. Standing near a cliff if we clap then, after sometime, we hear the reflection of the clap. This is due to the fact that the sound wave travel in all directions and when they come across another surface the sound waves undergo reflection and travel in different directions. As the picture of the object seen after reflection is called an *image*, the sound heard after reflection is called an *echo*. This can be demonstrated by a simple experiment as described below.

Take a long PVC pipe and cut it into two equal parts. Hold the two parts against a smooth surface. Place a watch at the open end of one tube and ask a student to place his ear against the open end of the second tube. Tell the student to slightly move the tube sideways till clear ticking of the watch is heard. Place a big cardboard sheet between the two tubes so that the sound does not reach the ear through any other path. Measure the angles that the two tubes make with the normal at the point of incidence as shown in Fig 12.21. Repeat the experiment by changing the angle of incidence. It will be found that in all cases the angle of reflection is equal to the angle of incidence.

The whispering gallery in the Shah Jehan Mosque Thatta provides an example of the reflection of sound. A sound whispered against

the wall on one side of the gallery can be heard clearly on the other side. Being circular in shape and made of stone, the walls reflect the sound of the whisper all round the gallery and concentrate the sound at the opposite side 32.6 m away. Normally a whisper would be inaudible at such a distance.

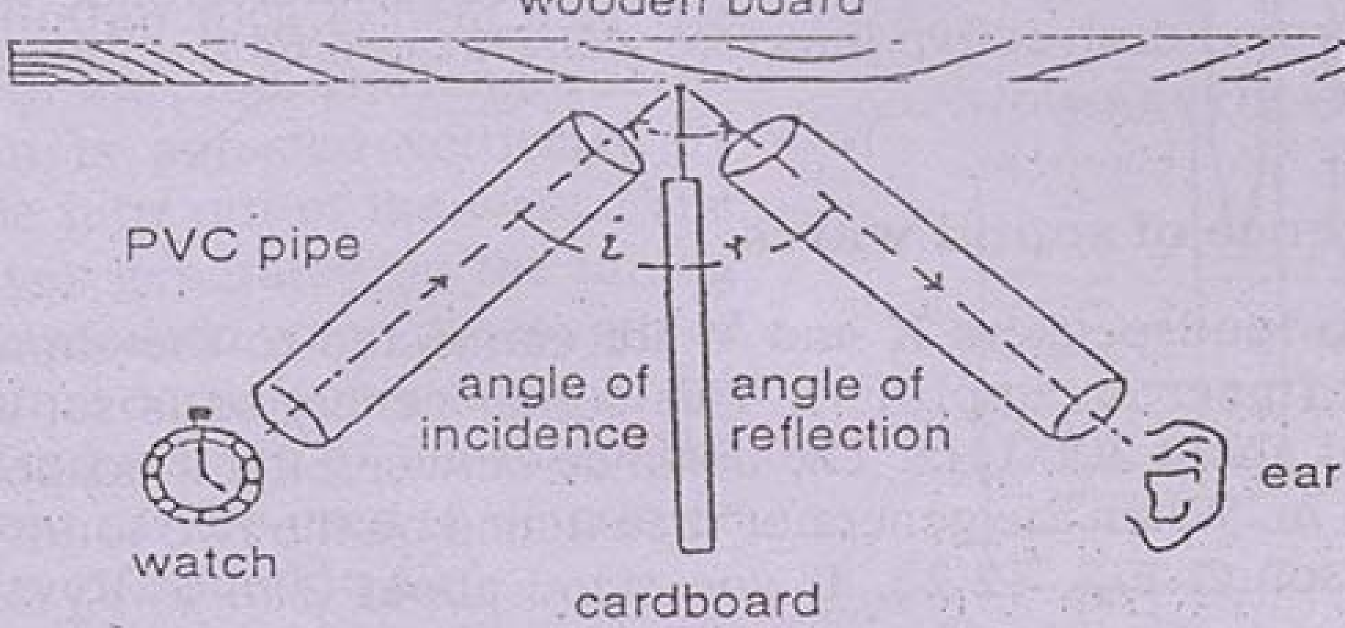


Fig. 12.21 Reflection from wooden board.

Another application of the reflection of sound is in whispering tubes and stethoscopes. A faint sound like that of the heart throb or the inhaling and exhaling air by the lungs is fed into a narrow flexible tube. This faint sound travels through the tube and reaches the ear drum after several reflections. Thus a sound ordinarily inaudible can be heard. A whispering tube is shown in Fig 12.22. A rolled up sheet of paper or cardboard tube can be used to test this out.

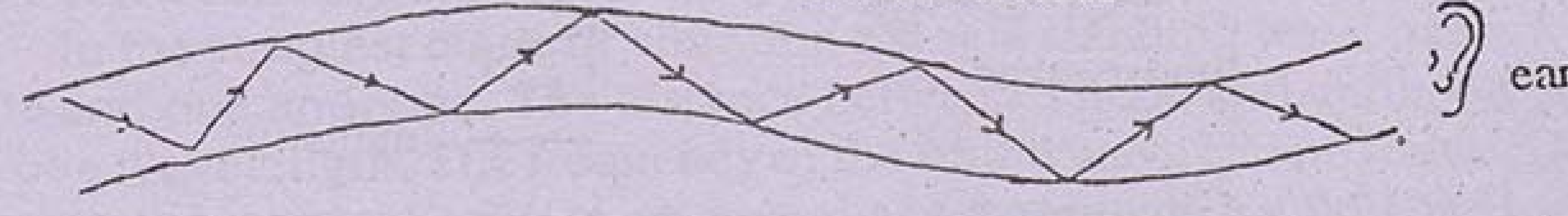


Fig. 12.22 Reflection of sound in a narrow tube.

12.15 INTERFERENCE OF SOUND

We have already discussed the interference of waves. As sound also travels in the form of waves, it must also exhibit interference. That is when two sound waves of the same frequency and amplitude pass through the same region of space at the same time, then at the place where the compression of one sound wave falls on the compression of

the second wave, we get an intense compression. Also we get an intense rarefaction if the rarefactions of the two waves coincide. The results in a louder sound. However, if the compression of one wave falls on the rarefaction of the second wave we hear no sound or a very faint sound. The former is called constructive interference while the later is termed destructive interference. This can be demonstrated by an activity given below.

Interference of sound waves

Two loudspeakers X and Y are connected to the same signal gererator (tuner or amplifier). This signal generator is set at a frequency of about 3000 Hz. The distance between the loudspeakers is about 0.5 m. When the generator is switched on the two loudspeakers produce sound Fig. 12.23. If you stand about 2 m away from the speakers and, after blocking one ear, move your head sideway through at least 0.2 metres you should hear variations in loudness.

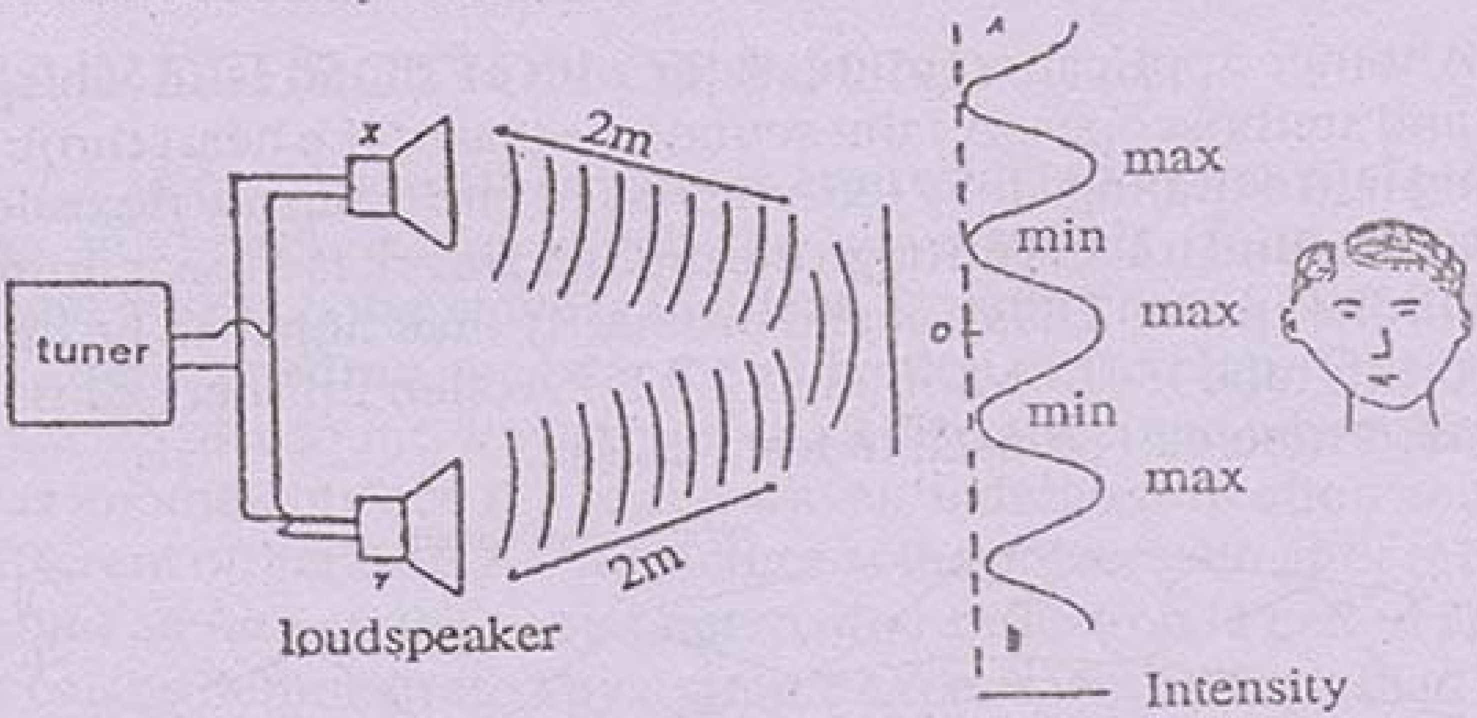


Fig. 12.23 Apparatus for studying the interference of sound.

A microphone may be used to detect the change in the intensity (loudness) of the sound by moving it along the line AB which is about 2 m away from the loudspeakers. The microphone detects the rise and fall of the loudness of the sound produced by the loudspeakers. This shows that interference of sound waves does take place.

Resonance in Sound Waves

We have already discussed the resonance phenomenon in waves in somewhat greater detail in the earlier article. Here we describe an

activity to show resonance in sound waves.

A simple apparatus for demonstrating resonance is shown in Fig. 12.24. A long verticle tube is partially dipped in water contained in a beaker. A vibrating tuning fork is held near the upper end of the tube. The length of the air column is adjusted vertically by moving the tube out of the water. The sound waves generated by the tuning fork are reinforced when the length of the air column corresponds to one of the resonant frequencies of the tube. The arrangement can be used to determine the velocity of sound in air.

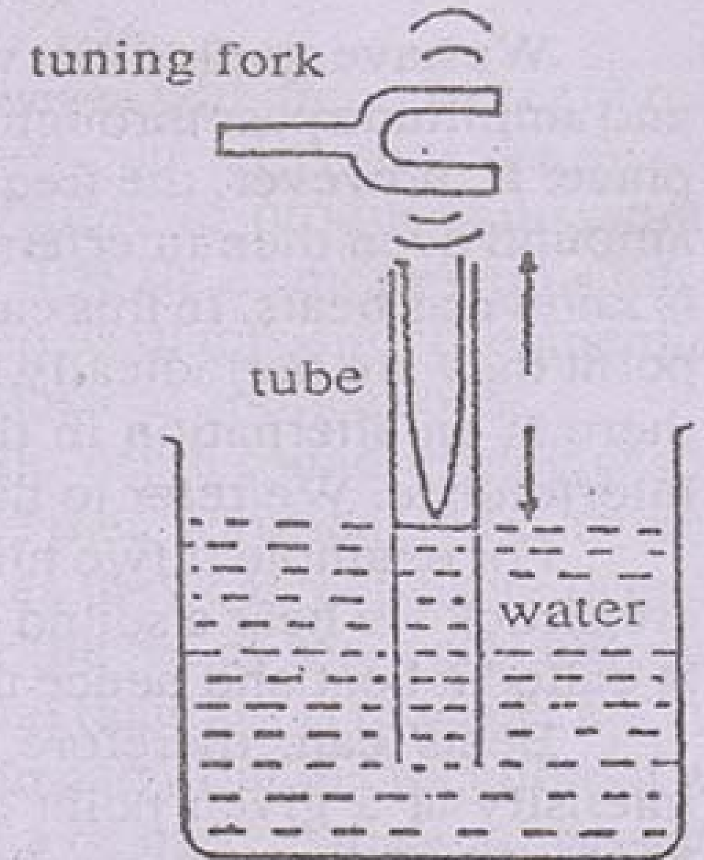


Fig. 12.24 Apparatus to demonstrate resonance

Whenever a sound wave comes across a barrier it is reflected back in the same medium. In this process, the reflected waves interact with the incident waves and produce stationary waves. When this happens we get a louder sound. The loud sound indicates that the reflected waves are in resonance with the incident waves produced by the tuning fork. In stationary waves those points where the disturbance is maximum are called antinodes. Two nodes or two antinodes can not exist consecutively but must be separated by an antinode or a node.

In the simplest mode of vibration a single standing wave has one node and one antinode. The distance between a node and antinode is 1/4 of a wavelength. The frequency of the sound wave can be found by moving the vertical tube in or out of the water until a position of maximum loudness is obtained. Measure from the water to the top of the tube. The reflection of sound waves at the upper end takes place a little distance above the open end. But this is usually ignored unless high accuracy is required.

The speed of sound can be experimentally calculated with the formula,

$$V = \lambda f$$

where the wavelength λ is four times the distance at which the maximum loudness is obtained.

Beats

We have studied that when two waves having the same frequency and amplitude pass through the same region of space interference takes place. If, however, the frequencies of the two waves differ by a small amound even then interference takes place. This type of interference is known as beats. In this case when two waves are observed at a given point they are periodically in or out of step with one another. That is, there is an alternation in time between constructive and destructive interference. We refer to this phenomenon as interference in time.

For example if two tuning forks of slightly different frequenceis areistruck we hear a sound of alternately high and low intensity. This is called a beat and hence the phenomenon is popularly called beats.

Beats can, therefore, be defined as the periodic variation in intensity at a given point due to the superimposition of two waves having slightly different frequencies. The number of beats one hears per second, or the beat frequency, is equal to the difference in frequency between the two sounds. The maximum beat frequency that a human ear can detect is 7 beats per second. A graphic representation of two waves and their super imposition is shown in Fig. 12.25.

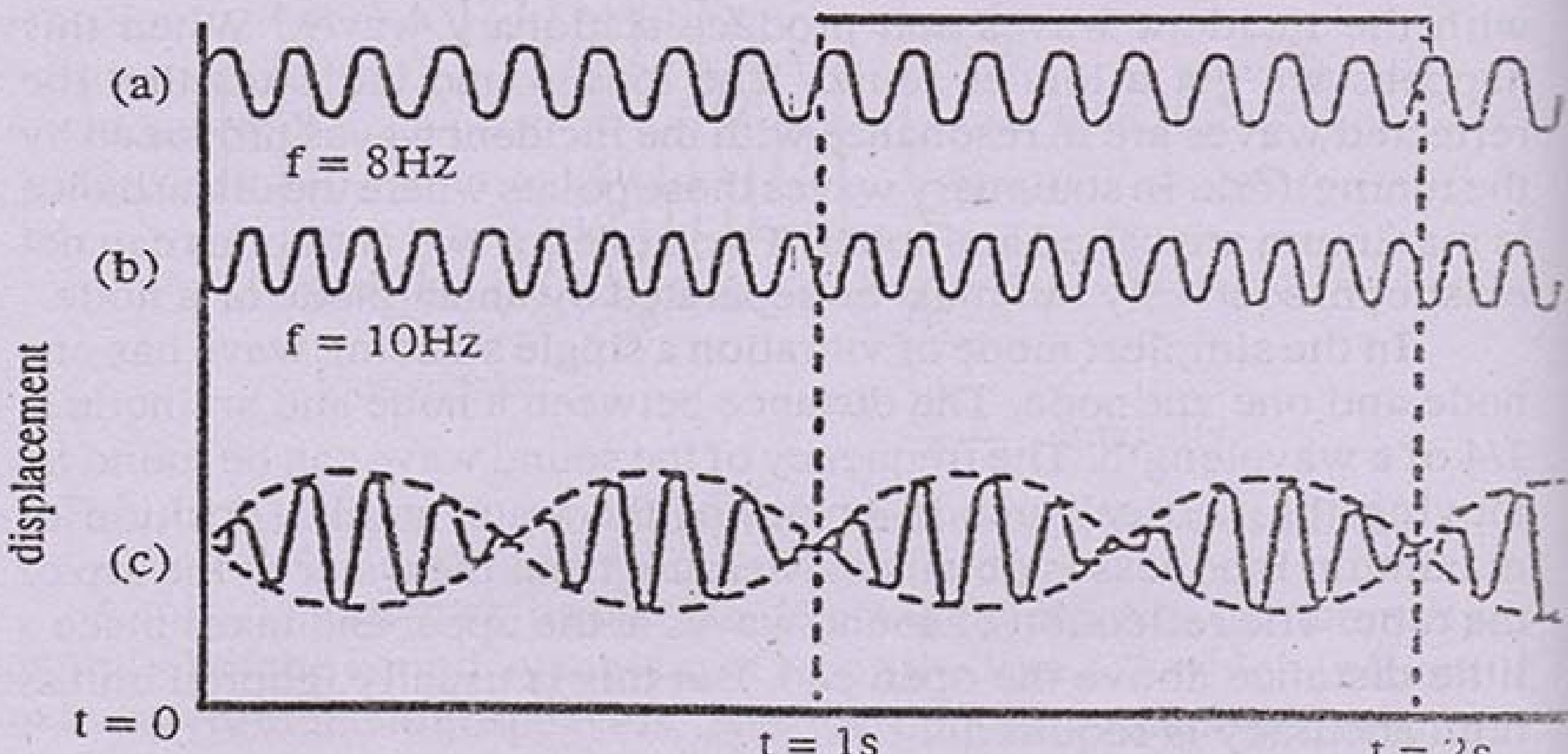


Fig. 12.25

(a) Wave of frequency 8 Hz. (b) Wave of frequency 10 Hz. (c) When (a) and (b) are super imposed the resultant waves shows 2 beats.

Ultrasonics

Ultrasonic waves are longitudinal waves with frequencies above the audible range. Such waves are usually produced by setting a quartz crystal to oscillate electrically. It is possible to produce ultrasonic waves of the frequencies of the order of 10^9 Hz or more with such a device. Ultrasonic waves are widely used as diagnostic, therapeutic, and surgical tools in medicine and in industrial applications.

Ultrasonic waves can be used in echo-depth sounding devices to determine the depth of the sea floor. Since their wavelengths are much shorter than those of normal sound waves they can penetrate deeper into the sea. Radar cannot be used under the sea as sea water absorbs microwaves. Sonar (Sound navigation and ranging) is used because it emits ultrasonic waves and can be used to carry out the location of an object by its echo. The same principle can be used to make ultrasonic guidance devices for the blind, to detect cracks in metal structures, to kill bacteria and micro-organism in liquids. These are also used to obtain cross-sectional pictures of patients in hospitals. Ultrasound scans are often preferred to X-ray scans, because ultrasound is much safer than X-rays. X-rays and ultrasound are used for different purpose in medicine. Ultrasound is considered best to examine the soft fleshy parts of the body i.e, examining the foetus in a pregnant woman, whereas X-rays are better for examining suspected broken bones which are denser than flesh. Ultrasound is also being used for cleaning places and objects which can not be cleaned in a normal way. Ultrasonic clearners are especially popular with jewellers and material scientists for cleaning delicate instruments and materials.

SUMMARY

— In executing a simple harmonic motion, the magnitude of acceleration of a vibrating body is always directly proportional to displacement from the mean position and direction of acceleration is always towards the mean position.

— Wave motion is that form of motion in which an oscillating disturbance is transmitted from one position to the next without the actual rectilinear translation of the particles of the medium. Waves are produced by certain kinds of disturbances or changes at certain places.

— Longitudinal waves are those waves in which the disturbance travels in the same direction in which the particles of the medium oscillate.
— Transverse waves are those waves in which the disturbance travels in a direction perpendicular to the direction of oscillation of the particles.
— Frquency is the number of waves generated by a source in one second, executed by an object that produces waves in one second.
— Time period or period is the time taken by a vibrating body to complete one vibration. It is denoted by T.

$$T = 1/f \qquad \text{or} \qquad f = 1/T.$$

— Wave Length is the distance covered by the wave during one period. It is also defined as the distance between two consecutive crests or between two consecutive troughs or it is the distance between two consecutive compressions or between two consecutive rarefactions. It is denoted by Greek letter lambda λ.
— The Amplitude of a wave is the maximum displacement of the particle from its mean position. It is denoted by A.
— The Wave Speed is the distance travelled by the wave in one second in the direction of wave. It is denoted by V.

$$\begin{aligned} V &= \lambda(m).f(s^{-1}) \\ &= f\lambda(ms^{-1}) \end{aligned}$$

— Waves whether they are transverse waves, or longitudinal waves are reflected whenever incident on a surface separating two media. Sound waves obey the same laws of reflection as are obeyed by light waves.
— When two waves of the same amplitude pass each other then the net amplitude of the combined wave will be the algebraic sum of the displacements of the two separate waves. If at a given point the crest (condensation) of one wave falls on the crest (condensation) of the other wave then the result is that a crest (condensation) larger than the crest of any of the two waves. This is called constructive interference. However, if the crest of one falls on the trough of the other wave the amplitude of the resultant wave is much smaller than the amplitude of any of the two waves. This is called destructive interference. This phenomenon is called interference.

— When two waves, having the same frequency and amplitude, travelling in the opposite direction meet one another the resulting interference pattern gives rise to stationary or standing waves. It is called a stationary wave because it does not appear moving.
— Resonance is the phenomenon in which an object whose natural time period happens to be the same as that of an other object which is vibrating in its vicinity begins to vibrate under the influence of these vibrations.
— Sound waves are longitudinal waves. When these waves fall upon an ear produce the sensation of sound.
— Waves are handed over from one layer of the medium to the next. This is known as propagation of sound. Sound waves can propagate through solids, liquids and gases.
— The velocity of sound is a maximum in solids and is a minimum in gases. The velocity of sound is about 340 ms^{-1} at room temperature in air.It is much smaller than the velocity of light which is 3×10^8 ms^{-1}
— The clap and echo method can be used to estimate the velocity of sound.
— Loudness is that characteristic of sound which enables us to distinguish between a loud and a faint sound. Loudness depends upon the amplitude and area of the vibrating body and the distance of the listener from the source. Loudness is a sensation.
— Pitch is that property of the sound which enables us to distinguish between a shrill and a flat sound. Pitch of a sound depends upon the frequency of the sound. The greater the frequency the higher the pitch.
— Quality or timbre is that characteristic of a sound by virtue of which we can distinguish between two sounds of the same pitch and loudness but produced by two different musical instruments. This is due to the presence of different overtones in the two sounds.
— Sound waves suffer reflection whenever they are incident upon a surface separating two media. Sound waves on reflection from a surface obey the laws of reflection. (i) Angle of incidence equals to angle of reflection (ii) The incident sound, the reflected sound and the normal at the point of incidence all lie in the same plane.
— The sound heard after reflection from a surface is called an echo.
— When two sound waves that differ from one another in frequency by a very small number are heard together then sometimes we hear a loud sound where the compressions of the two waves meet and at other time we hear a faint sound where the compression of one wave falls upon the rarefaction of the other wave. This phenomenon is called beats. The number of times we hear the rise and fall of sound per second

is called the beat frequency. The beat frequency is equal to the difference in the frequencies of the two sounds.

— The range of frequency of the waves with in which a normal healthy human ear can detect a sound is called the audible frequency range. It varies between 20-20,000 Hz. A normal human ear can not hear a sound whose frequency is less than 20 Hz or more than 20,000 Hz.

— Ultrasonic wayes are longitudinal waves with frequencies above the audible range. Such waves are usually produced by making a quartz crystal oscillate electrically.

QUESTIONS

12.1 Write answers to the questions given below:

(i) Describe the phenomenon of waves with the help of experiment.

(ii) Prove that the vibratory motion oı a mass attached to a spring is simple harmonic motion

(iii) Explain the difference between transverse and longitudinal waves.

(iv) Define the terms wavelength, frequency and time period. Derive an equation between velocity, wavelength and frequency.

(v) How is the sound produced. What is the range of frequencies of audiable sound.

(vi) What is resonance and explain it in some detail.

(vii) What is the difference between musical sound and noise.

(viii) Describe the characteristic of musical sound.

(ix) What are the uses of ultrasound.

(x) Why do the explosive sounds produced in the sun not heard on the earth?

(xi) Why does the flash of lightning seen earlier than the sound of thunder.

12.2 Fill in the blanks.

(i) A motion which repeats itself in equal intervals of time is called________.

(ii) _______ is also a periodic motion.

(iii) While executing a simple harmonic motion, the magnitude of the acceleration of the body is ______ to its distance from the mean position and the direction of the acceleration is always towards the _______.

(iv) The vibratory motion of the bob of a simple pendulum is_____.

(v) The waves in which the particles of the medium vibrate in a direction perpendicular to the direction of propagation of waves are called ________.

(vi) Transverse wave consist of ________.

(vii) Longitudinal waves consist of _____ and _______ and as they pass through a medium, the particles of the medium vibrates in the _____ of the waves.

(viii) Wave length is the distance between two such _______ particle which are in the similar _______ of these periodic motion.

(ix) Human ear cannot hear sound of frequency greater than_______ and it is known as______.

(x) An interesting example of resonance is that of _________.

(xi) Sound which is agreeable to human ear is called _____ sound.

(xii) The characteristics of musical sound are (a)______ (b)_______ (c)_______.

12.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) A body with frequency f would complete one vibration in ____ second.

$$(a)\frac{1}{f} \quad (b)\frac{1}{f^2} \quad (c)1 \quad (d)f$$

(ii) If the length of the pendulum becomes four times, it's time period will become _______.
(a) four times (b) twice (c) three time (d) eight times

(iii) In transverse waves the distance between two consecutive crests or between two consecutive troughs is called________.
(a) displacement (b) wave length (c) velocity (d) speed

(iv) If the mass attached to a spring becomes four times, the time period of its S.H.M. will become_______.

(a) twice (b) three times (c) four times (d) eight times

(v) If frequency of waves f = 30 cycles per second and wave length $\lambda = 0.2$ metre, then the velocity of waves is _____ per second.

(a) 6 (b) 150 (c) 0.0066 (d) 8

(vi) The waves produced by a vibrating body in air are _______ waves.

(a) longitudinal (b) transverse (c) electromagnetic (d) magnetic

(vii) Speed of sound in air at normal temperature pressure (N.T.P) is _______ m/s

(a) 336 (b)672 (c) 712 (d) 785

12.4 Pick out true and false.

(i) Amplitude is the distance between the extreme positions of a vibrating body.

(ii) The time required to complete one vibration is called time period.

(iii) The motion of the bob of a simple pendulum is not S.H.M.

(iv) Force applied to a spring is inversely proportional to the extension in the spring.

(v) As the waves pass over the surface of water, the particles of water start moving with the waves.

(vi) The velocity of a wave is equal to the product of the frequency and wave length of the wave.

(vii) Wave length is the distance between two such consecutive particles of the medium which are in the similar state of their periodic motion.

(viii) Waves is the only effective way of transferring energy from one place to another.

(ix) The pitch of the sound increases if the frequency of the vibrating body is increased.

(x) Generally pitch of the sound of women is lower than the pitch of the sound of men.

(xi) The characteristics of sound which helps us to distinguish the soundness of the same pitch and loudness is called the quality of sound.

(xii) The frequency of a wave is directly proportional to its time period.

PROBLEMS

12.1 The wave length of a wave is 0.1 nm. Its speed is 3×10^8 m s^{-1}. What is the frequency of the wave.

(3×10^{18} Hz.)

12.2 A tuning fork vibrates 256 times each second and produces a wave 1.3 m long. Calculate (a) the period and (b) the velocity. of the wave.

(3.9×10^{-3} s, 332.8 m/s.)

12.3 A radio station broad casts an AM radio waves whose frequency is 1230×10^3 Hz and an FM radio wave whose frequency is 91.9×10^6 Hz. Find the distance between adjacent crest in each wave.

(The speed of AM and FM radio wave is 3.00×10^{10}cm/s)

(24390 cm,326.44cm)

12.4 Find the time period of a simple pendulum whose length is 144 cm.

(2.41 s)

12.5 A body of mass 0.3 kg is attached to a horizontal spring. If the value of the spring constant is 5 N/m, find the time period of the body if it is given a small displacement.

(1.54 s)

12.6 A piece of paper completes 50 vibrations in 5 s when some waves pass through the surface of water. Find the time period and the frequency of the piece of paper. If the wave-length of the wave is 10 cm, find the velocity of the waves.

(0.1 s, 10Hz, 100 cm/s)

12.7 40 waves pass through a point on the surface of a pond in 2s. Calculate the wave-length if the velocity of waves is 3.5 m/s.

(0.175 m)

12.8 Calculate the length of a second's pendulum taking g equal to $9.8 m/s^2$ (A second's pendulum is a simple pendulum having a time period of 2 seconds).

(.994 m)

CHAPTER - 13

PROPAGATION AND REFLECTION OF LIGHT

13.1 INTRODUCTION

When we see blue sky, shining stars, rising sun, deap sea, green plants, and colourful flowers, we enjoy and appreciate the beauty of the universe and our earth. We enjoy these phenomenon due to the presence of light. When we say that we see an object, rays of light from it enters our eyes and evokes the sensation of vision. In the above example we find that there are objects that emit light and hence are seen There are other objects that are seen because they reflect light into our eyes. In order lo understand light and its nature it is necessary to study some of its fundamenlal properties. These are linear propagation, reflection, refraction, dispersion and emission of light.

13.2 RECTILINEAR PROPAGATION OF LIGHT (PATH OF LIGHT)

We commonly observe that when light coming from a source falls on an opaque object a shadow is formed behind it. This property of light lead us to the idea that light travels in a straight line.

13.3 THE PINHOLE CAMERA

A pinhole camera is one of the practical applications of rectilinear propagation of light. It was invented in the sixteenth century to observe eclipses of the sun without damaging the eyesight it is also known as *camera obscura* (dark room). It consists of a rectangular box containing a very small hole on one side and a frosted glass plate, tracing paper or a photographic film on the opposite side. When a narrow pencil of rays of light starting from an object is allowed to pass through the pinhole, an inverted and real image of the object is formed on the back of the camera. For observing a clear image, external light is excluded by covering the box wilh a dark cloth.

In order to understand the image formation of the pinhole camera consider the following diagram as shown in Fig 13.1

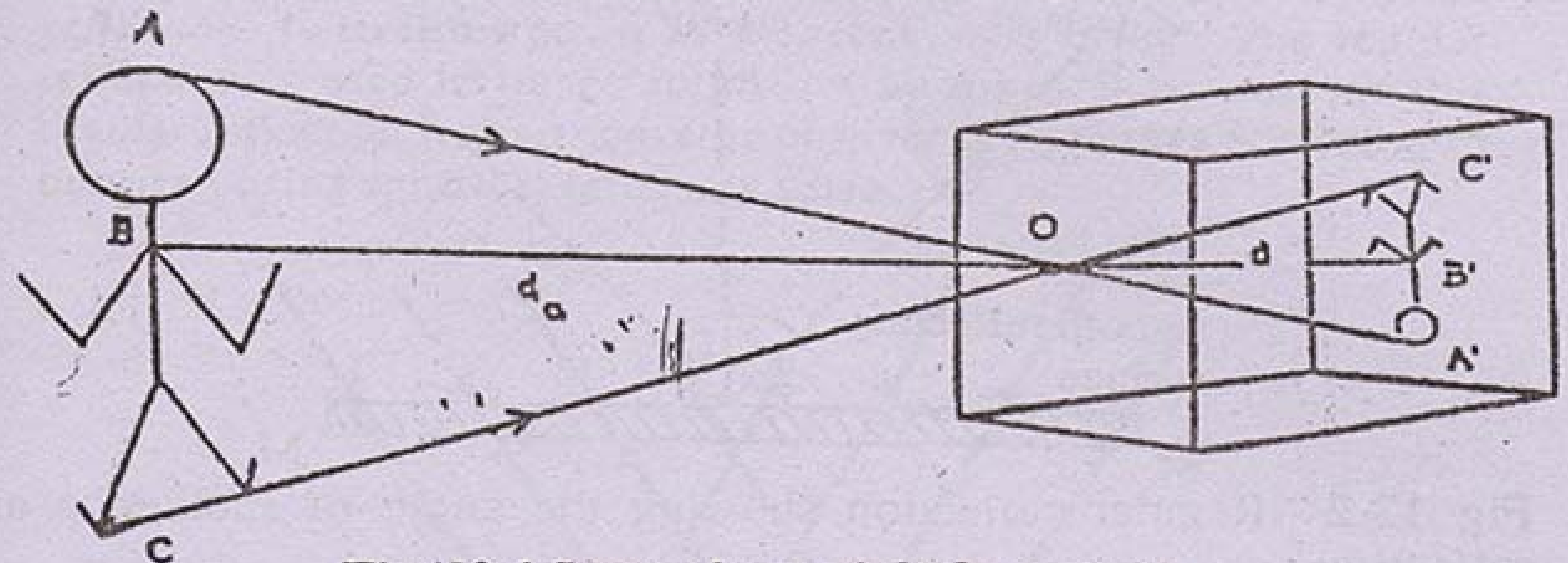


Fig. 13.1 Image by a pinhole camera

A narrow pencil of rays starting from point A passes through the pinhole O and, illuminates a small area at A′. Similarly a narrow pencil of rays starting from C illuminates a small area at C′. In this way points lying between A and C illuminate corresponding points between A′ and C′ and a real and inverted image A′ C′ of the object AC is formed on the back of the camera.

13.4 REFLECTION OF LIGHT

We know that most of the objects which we observe do not emit light. In our solar system, the sun is the only object which emits light yet we observe other planets and satellites e.g, mercury and the moon. When light spreading from a source in one medium strikes the surface of another medium a part of it is sent back in the same medium or reflected, a part is absorbed and a part is transmitted into the other medium (or refracted) if it is transparent or translucent. The phenomenon of reflection is widely used in our life. For example the reflection of radar waves is used for safe take off and landing of aeroplanes.

13.5 LAWS OF REFLECTION

Light travels in a straight line in a homogeneous medium. When light travelling in one medium strikes the surface of another medium a part of it is reflected in the same medium. Mirrors and highly polished opaque surfaces reflect light strongly and the images that are formed due to this characteristics of reflection follow certain laws. These are discussed below

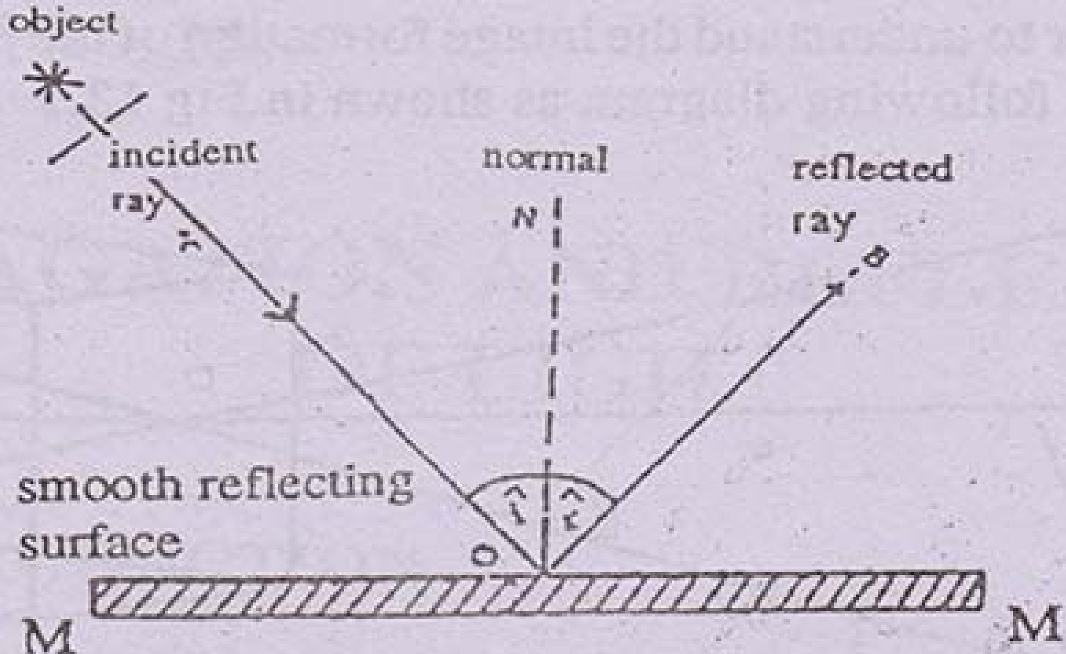


Fig 13.2 Regular reflection showing the angle of incidence and reflection

Fig 13.2; shows an arrangement consisting of a luminous object placed in front of a plane mirror. A ray of light AO is incident obliquely on the surface of the reflecting plane mirror MM at a point O, the point of incidence.

The incident ray is reflected from the reflecting surface so the ray OB is called the *reflected ray*. The angles 'i' and 'r' are the angles which the incident and reflected ray make with ON, the normal or perpendicular drawn to MM at the point of incidence 0. The angles AON and NOB so formed are called angle of incidence 'i' and *angle of reflection* 'r' respectively. These angles are found to be the same in magnitude. We also observe that the incident ray, reflected ray and the normal lie in the same plane satisfying the following conditions or laws, (Fig 13.2).

1. The angle of incidence is equal to the angle of reflection.

   Mathematically, we can express this as follows.

   i.e. $m\angle\hat{i} = m\angle\hat{r}$

2. The incident ray, the reflected ray and the normal all lie in the same plane.

These are called the *laws of reflection*.

13.6 REGULAR AND IRREGULAR REFLECTION

We observe that when parallel rays of light strike a smooth and shining surface, like a plane mirror, most of them are reflected in a

particular direction (Fig. 13.3). Such a reflection is called a *regular reflection*. It is also known as a *specular reflection*. The regular reflection is used in image formation by mirrors. Because of this regular reflection we can converge or diverge light rays according to our need using spherical reflecting surfaces.

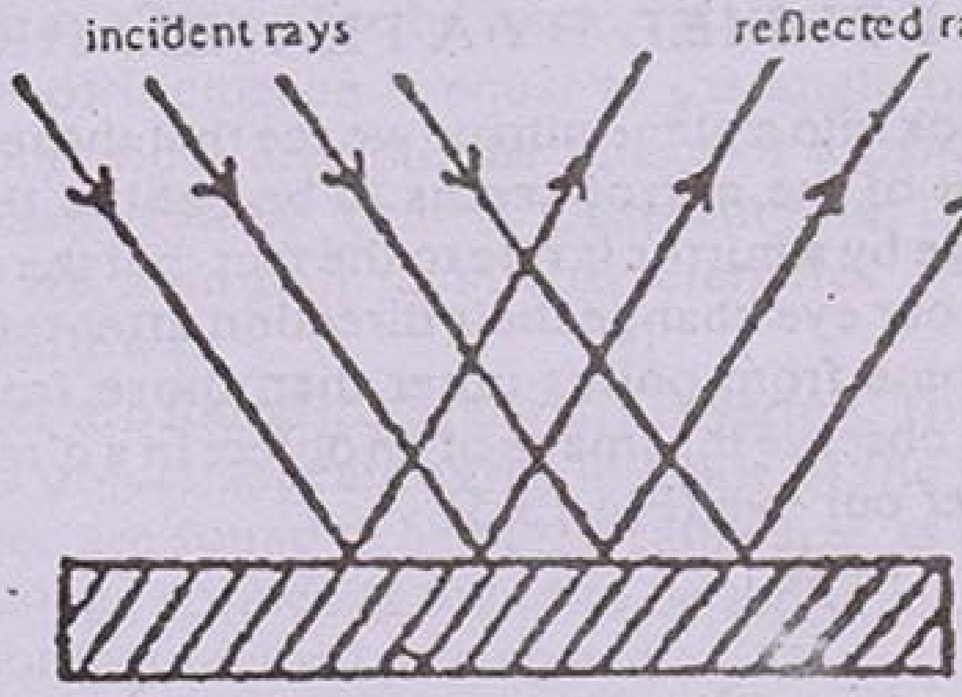


Fig. 13.3. Regular reflection

When a parallel beam of light is incident on surface's like white paper, or a painted wall, the reflected beam is scattered in different directions (Fig. 13.4). The reason for this random scattering is due to the highly irregular nature of these surfaces, which can be observed by using a microscope. Due to the roughness of the surface the angle of incidence is not the same for each ray, therefore, the reflected rays scatter in different directions. This type of reflection is known as *diffuse* or *irregular reflection*.

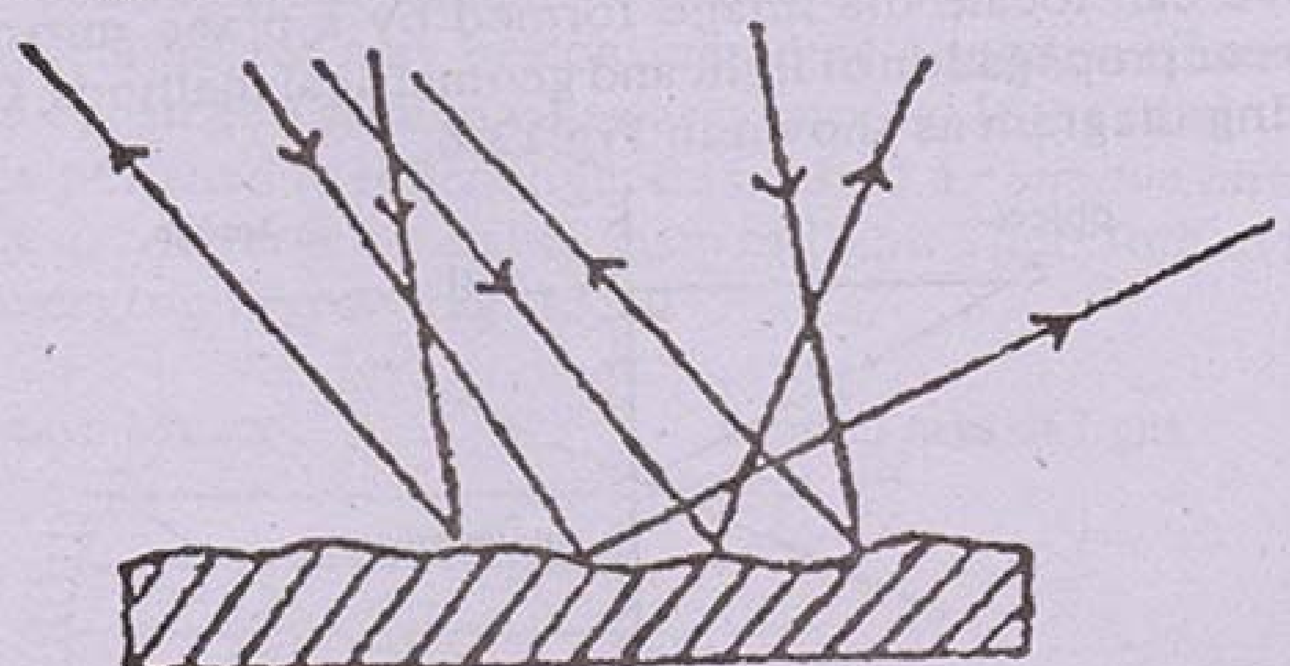


Fig. 13.4 Irregular reflectiion

Irregular reflection is very important for us. The gradual change of light, by which we receive light at dawn or after sunset, is due to irregular reflection of sunlight from the dust particles in air. All the non-luminous objects are visible due to the irregular reflection of light from their surfaces.

13.7 IMAGE FORMED BY A PLANE MIRROR

When we look into a plane mirror we see that those objects which are really in front of the mirror seems to be behind the mirror. The formation of image by a mirror is due to the fact that the rays travelling from an object to our eye change their direction after reflection so that they appear to come from points other than those from which they really started. We observe the image of an object in a direction in which the light rays enter our eye (Fig 13.5).

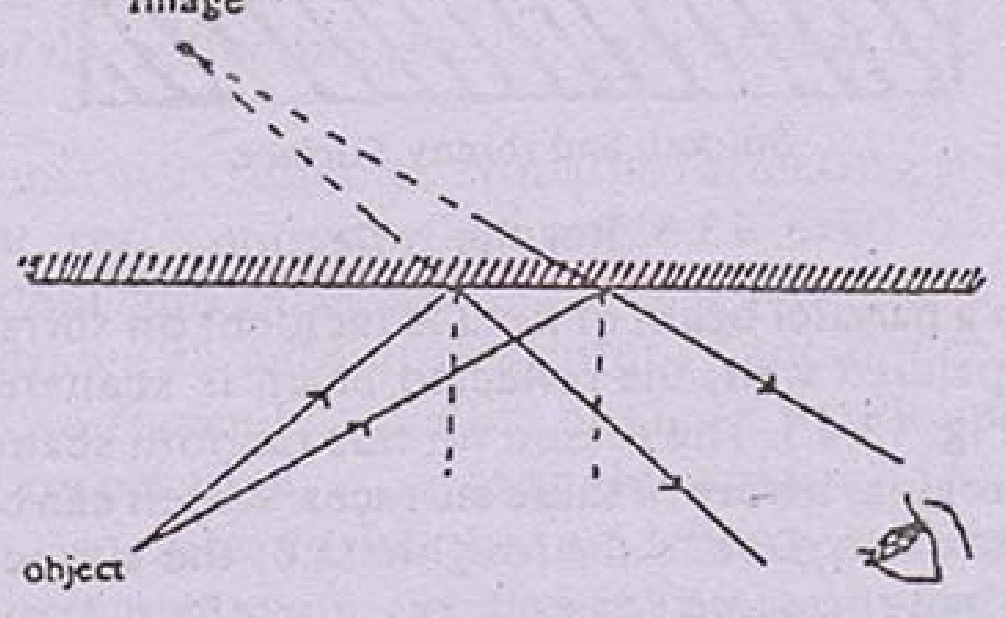


Fig. 13.5. Image formed by a plane mirror

We can locate the image formed by a plane mirror using the rectilinear propagation of light and geometrical methods. Consider the following diagram as shown in Fig.13.6.

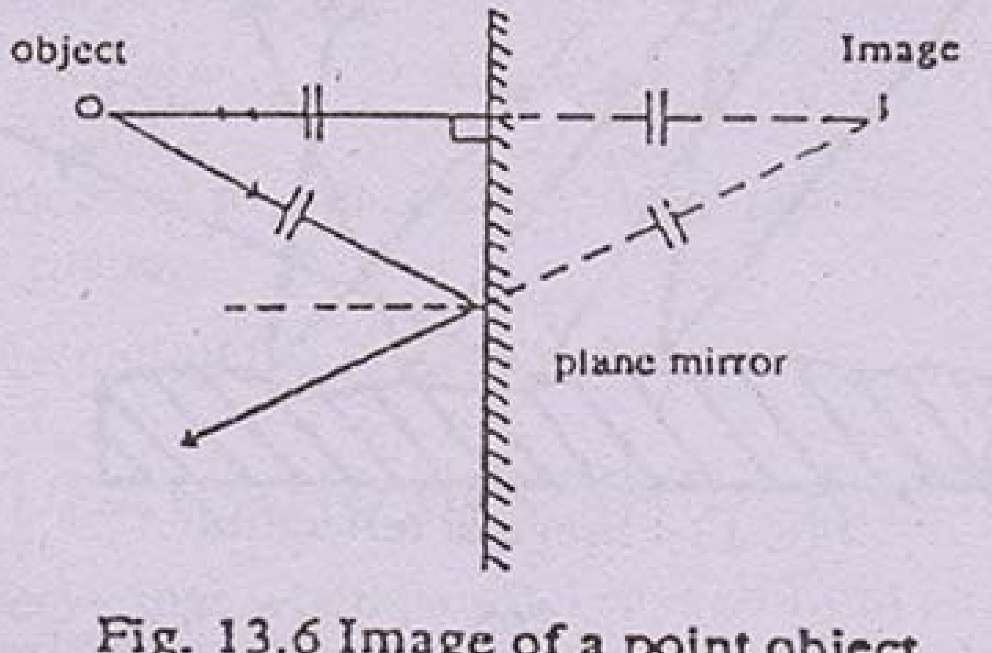


Fig. 13.6 Image of a point object

Light rays coming from a luminous object O are reflected by the plane mirror M and enters our eye. The line which joins the image I and the object O makes an angle of 90° with the surface of the mirror M. From the geometrical construction, the distance OM and IM are equal. Therefore, we feel that the light rays come from I but in fact they come from O and are reflected by the mirror. At the mirror surface we find that the angle of incidence is equal to the angle of reflection. Using the same geometrical method, the image of a candle placed in front of a plane mirror can be located. Observing the images formed by a plane mirror we note four main characteristics of the images which are given as follows:

1. Images are found to be laterally inverted. That is, the right side of the object appears as the left side of the image.
2. Images are found to be of the same size as that of the object.
3. The image formed is found to be errect and virtual, that is, it can not be obtained on a screen.
4. The image is as far behind the mirror as the object is in front of the mirror.

13.8 SPHERICAL MIRRORS

Curved or spherical mirrors are used for several purposes, we enjoy in fun houses observing our distorted images formed by curved mirrors. They also help us to avoid accidents at sharp turns in hilly areas. Safe driving is not possible without them on a busy road. Astronomers use large spherical mirrors in telescope to concentrate the rays of light coming from distant stars and planets.

A spherical mirror may be considered as a section of a hollow sphere. Now consider a section taken as a sample. If the inner side of the surface is polished to reflect light it is called a *concave mirror*. It has an ability to converge a parallel beam of light. Therefore, it is also called a converging mirror (Fig. 13.7).

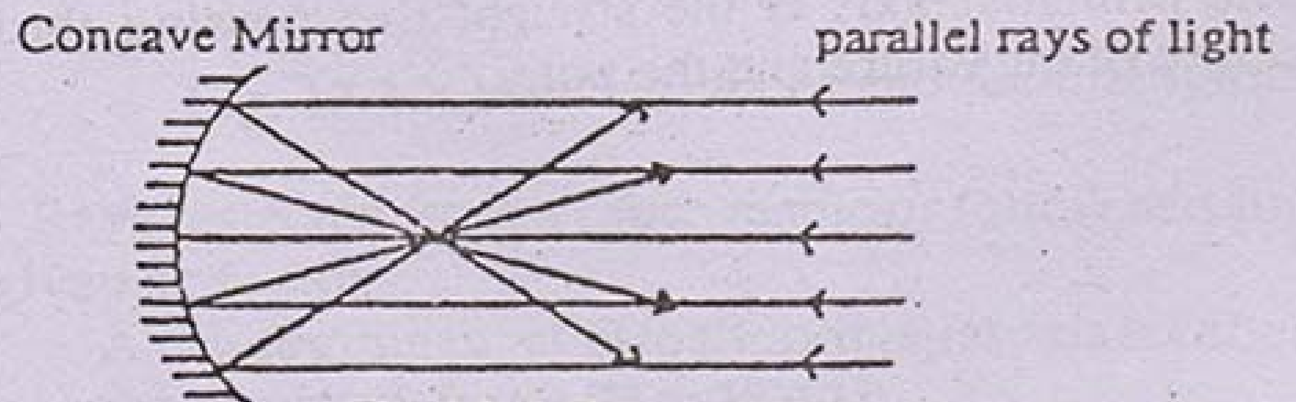


Fig.13.7 Converging or concave mirror.

If the outer side of the surface of the section is polished to reflect light it is called a *convex mirror*, it has an ability to diverge the parallel beam of light which is incident on it. Therefore, it is also known as a diverging mirror. (Fig.13.8)

At every point in both types of spherical mirrors reflection takes place in accordance with the laws of reflection. Therefore every small section of a spherical mirror behaves like a plane mirror.

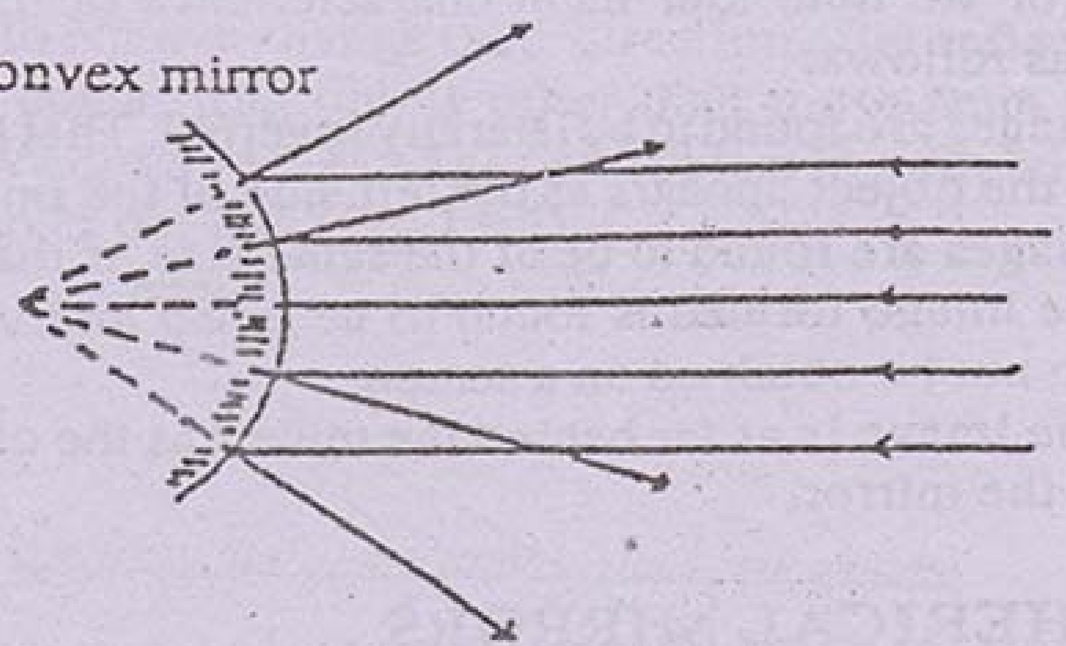


Fig.13.8 Diverging or convex mirror.

In order to explain the formation of images by spherical mirrors and their behaviour we define the following terms (Fig.13. 11).

(i) The centre of curvature (C) is the centre of the sphere from which a curved reflecting surface is obtained.

(ii) Radius of curvature is a straight line drawn from the centre of curvature to the reflecting curved surface.

(iii) The pole or vertex is the geometric centre of a curved mirror.

(iv) The principal axis is the straight line passing through the centre of curvature and the pole.

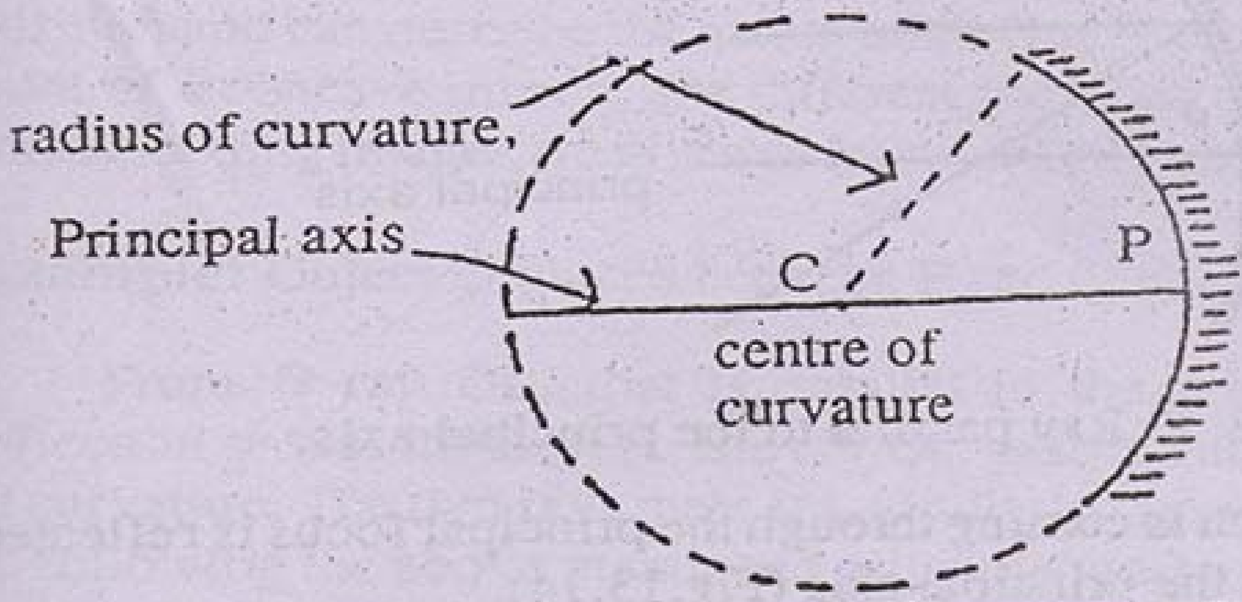


Fig.13.11 Principal axis, pole, centre of curvature and radius of curvature.

13.9 REFLECTION AND FORMATION OF IMAGES BY A CONCAVE MIRROR

When rays parallel to the principal axis, coming from a distant source or an object, strike a concave mirror, after reflection all the rays converge to the same point on the principal axis. This point is known as the *principal focus* (F). The distance between the principal focus and pole of the mirror is known as its *focal length* (f) (Fig. 13.12).

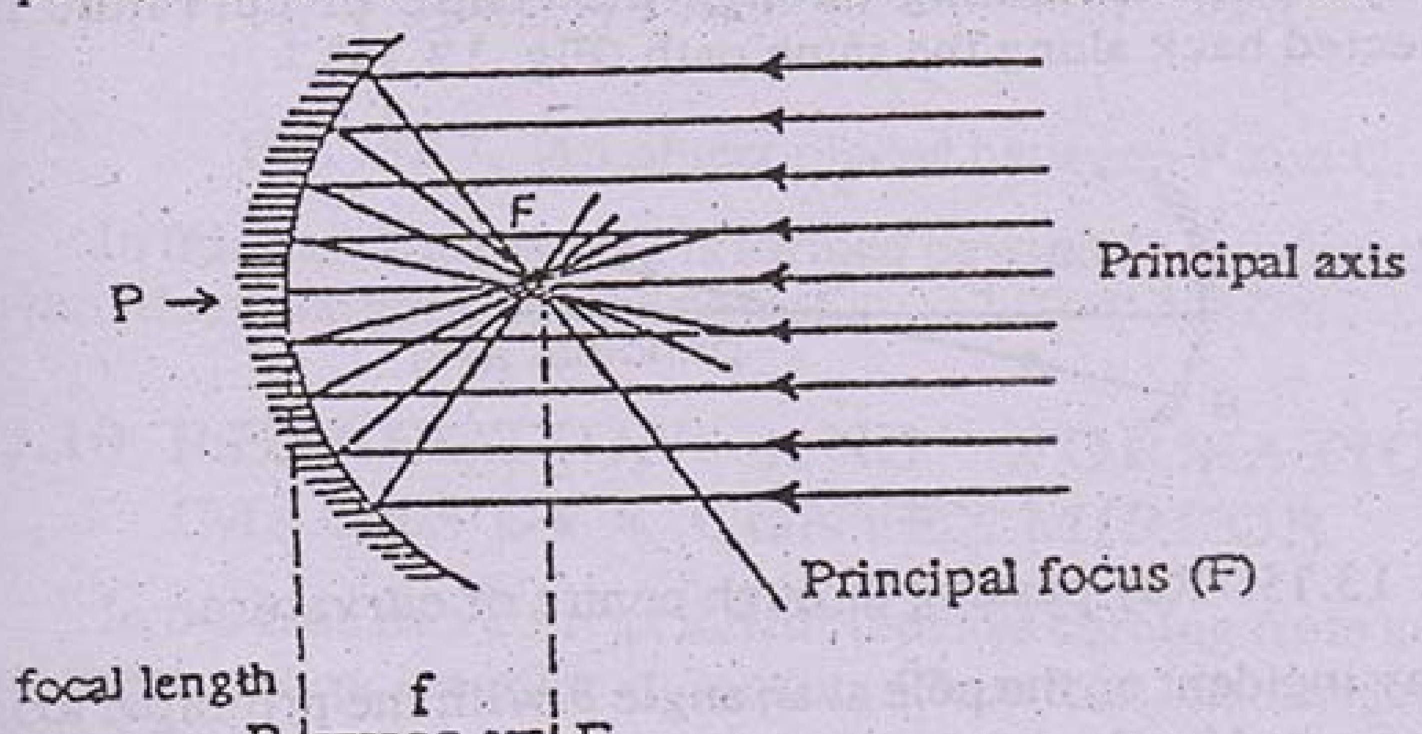


Fig. 13.12. Principal focus and focal length.

Rays reflected from a concave mirror follow certain rules which are given as follows.

1. A ray coming parallel to the principal axis is reflected through the principal focus (Fig. 13.13).

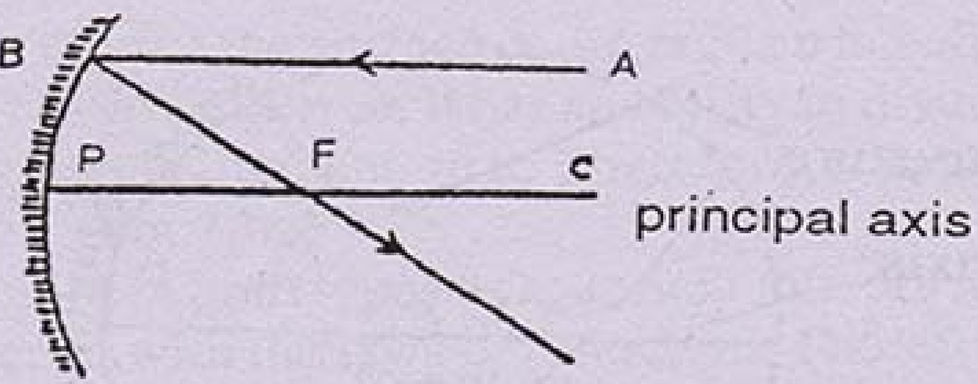


Fig. 13. 13 Ray parallel to tbe principal axis.

2. A ray which is coming through the principal focus is reflected parallel to the principal axis (Fig.13.14).

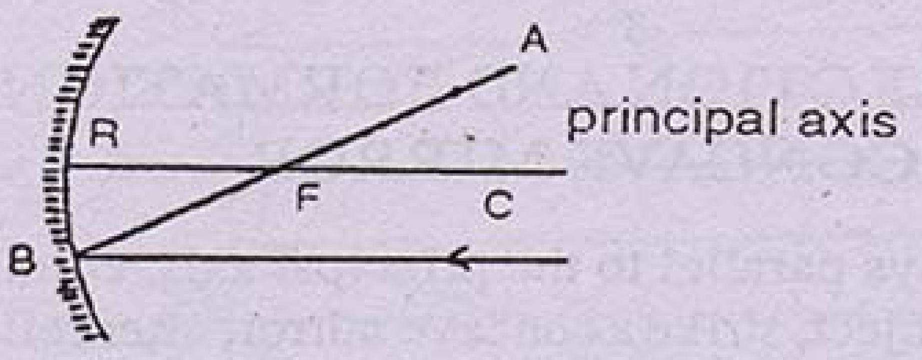


Fig. 13.14. Ray passing through the principal focus.

3. A ray which is coming through the centre of curvature is reflected back along the same path (Fig. 13.15).

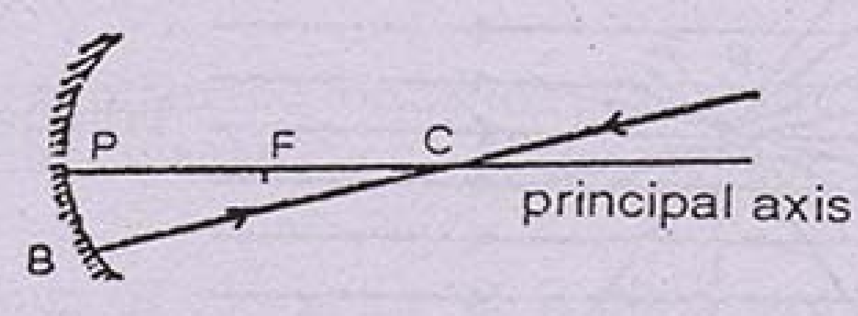


Fig. 13.15. Ray passing through centre of curvature.

4. A ray incident on the pole at an angle θ with the principal axis is reflected back at the same angle (Fig. 13.16).

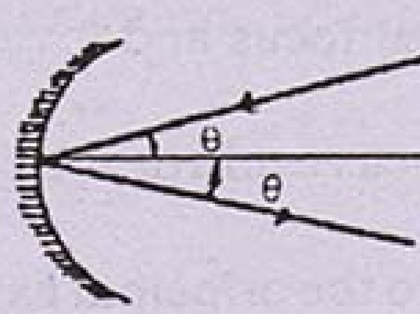


Fig. 13.16. A ray incident on the Pole at an angle θ

With the help of any of the rays mentioned above we can locate the position and can describe the nature of the image of an object placed in front of a concave mirror. The different positions of an object and its corresponding images are given as follows: Only one example is given.

Example: Object O placed between F and C.

From O ray OL, that is parallel to the principal axis, after reflection passes through F. The ray OC passes through C, the centre of curvature. The two rays meet at I. Similarly rays emitted from point A, following the above mentioned principle meet and form the image B. The image of AO is therefore BI after reflection.

Consider the ray diagram as shown in Fig. 13.17.

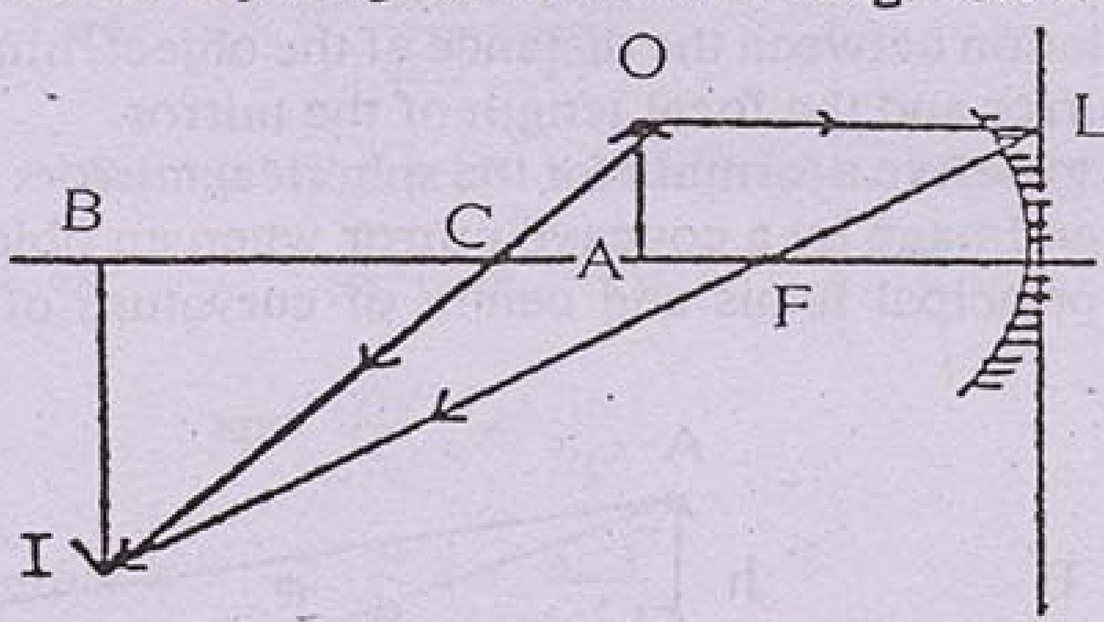


Fig. 13.17. An object placed between F and C.

In this case the image I is formed beyond C. It is found to be real, inverted and larger than the object.

13.10 REFLECTION AND FORMATION OF IMAGES BY A CONVEX MIRROR

In the case of a convex mirror the rays coming from an object do not meet at a point after reflection. Therefore, an image is formed behind the mirror shown by extending the reflected rays backward as dotted lines.

In a concave mirror real as well as virtual images are formed but in the case of a convex mirror only virtual images are produced. All these images are located between the pole of the mirror and the principal focus. They are found to be erect, virtual and smaller in size than the object (Fig.13.18).

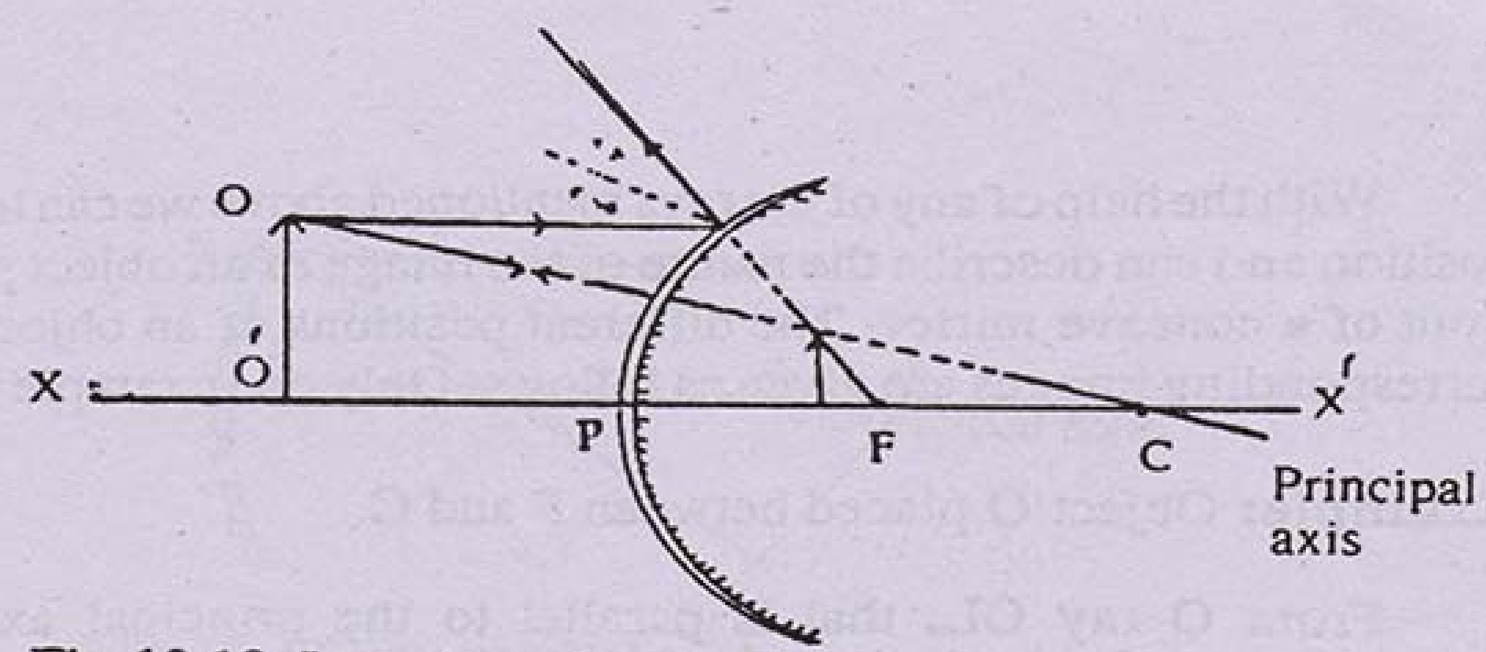


Fig.13.18 Location of image formed by a convex mirror

13.11 MIRROR FORMULA (EQUATION FOR SPHERICAL MIRROR)

It is a relation between the distance of the object, image from the pole of the mirror and the focal length of the mirror.

In order to derive a formula for the spherical mirrors consider the formation of an image by a concave mirror when an object is placed between the principal focus and centre of curvature of the mirror. (Fig. 13.19).

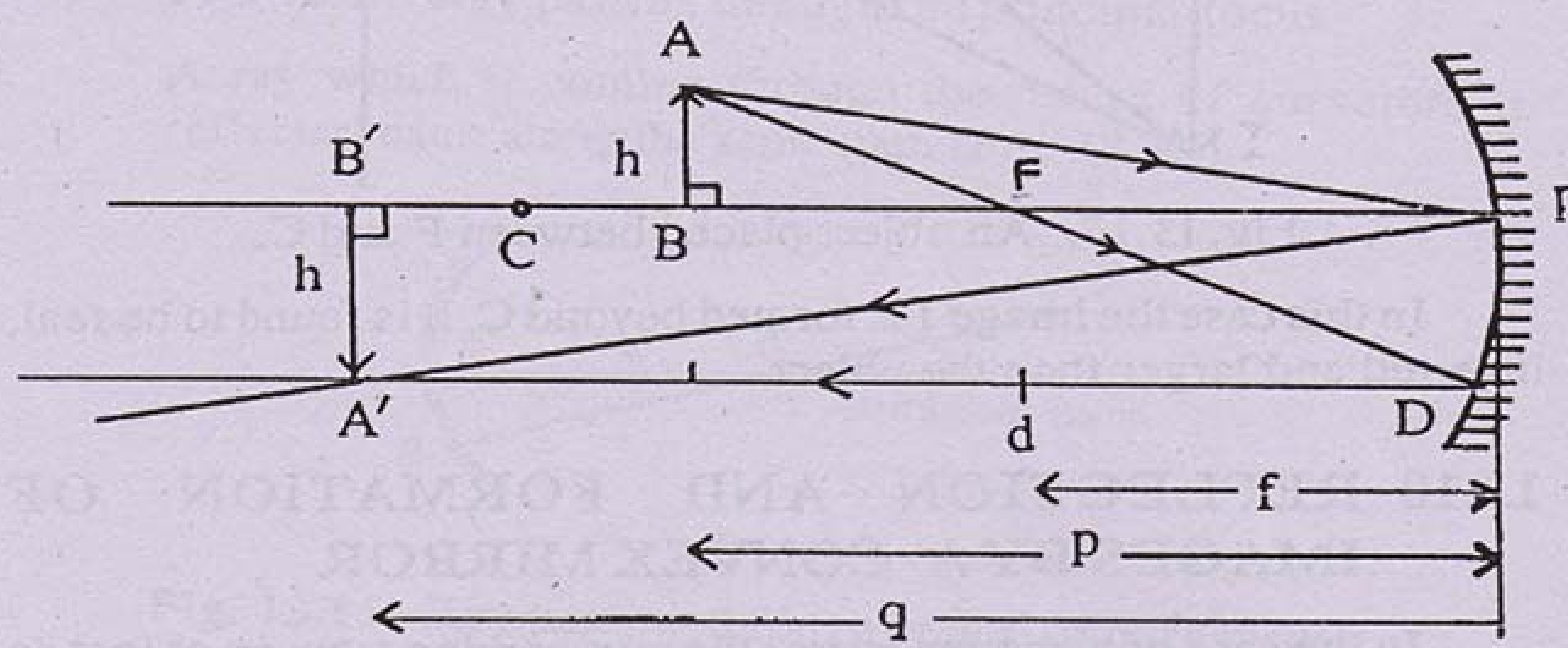


Fig.13.19. Focal length, object distance, image distance, and object and image height.

AB is an object placed before the mirror between the focus and the centre of curvature. Two rays AP and AD are incident on the mirror.

Ray AP is reflected with the same angle along the direction PA obeying the law of reflection.

As triangles A′PB′ and APB are similar.

Therefore,
$$\frac{AB}{A'B'} = \frac{PB}{PB'}$$

or
$$\frac{h_o}{h_i} = \frac{p}{q}$$

Ray AD which passes through F becomes parallel to the principal axis PB.

As triangles ABF and FPD are similar.

Therefore,
$$\frac{AB}{DB} = \frac{BF}{FP}$$

As $AB = h_o, DP = A'B'$ and $A'B' = h_i, FP = f$ and $BF = P - f$

but as
$$\frac{h_o}{h_i} = \frac{p-f}{f}$$

and
$$\frac{h_o}{h_i} = \frac{p}{q}$$

Therefore
$$\frac{p}{q} = \frac{p-f}{f}$$

Dividing both sides by p we get

$$\frac{p}{pq} = \frac{p-f}{pf}$$

$$\frac{1}{q} = \frac{1}{f} - \frac{1}{p}$$

Re-arranging the above equation we get

$$\frac{1}{f} = \frac{1}{p} + \frac{1}{q}$$

The above equation is known as the mirror equation or mirror formula.

13.12 SIGN CONVENTION AND MAGNIFICATION FORMULA

In the case of spherical mirrors virtual as well as real images are formed which under certain conditions are diminished and under other

conditions are magnified. Therefore, it is necessary to define a sign convention so that we may be able to distinguish between real and virtual images as well as enlarged and diminished images. The sign convention is given as follows:

(i) All distances are measured from the pole of the mirror.

(ii) Distances of real objects and images are taken as positive.

(iii) Distances of virtual objects and images are taken as negative.

Magnification: The ratio between the image height to object height is called magnification.

Using these conventions as given above, for an image formed by a concave mirror when an object is placed at a distance greater than the focal length, q and p both are positive.

Therefore $$M = \frac{h_i}{h_o} = \frac{q}{p}$$

13.13 USES OF SPHERICAL MIRRORS

There are several important uses of spherical mirrors some of them are given as follows:

(i) A concave mirror is used in a microscope to illuminate the object. In a telescope it is used to concentrate the parallel beams of light coming from distant stars.

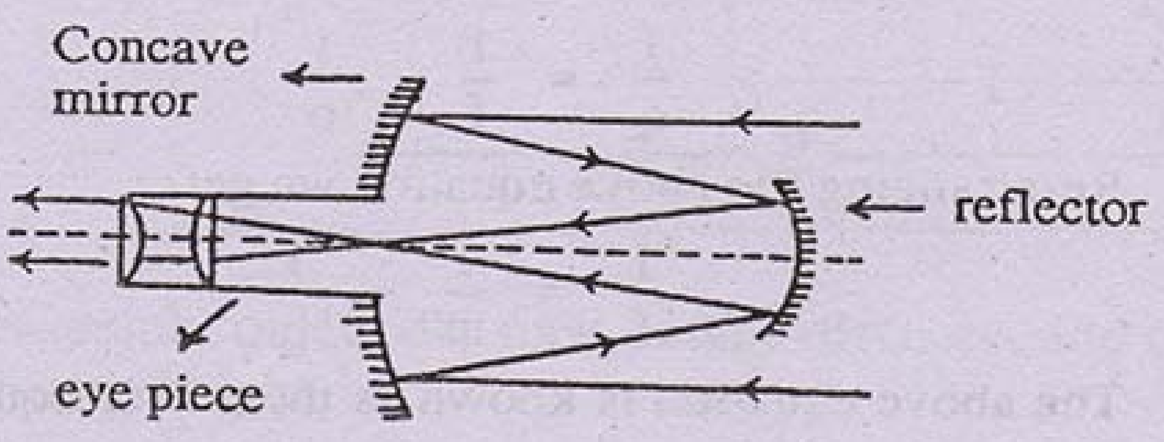


Fig. 13.20. Reflecting telescope.

(ii) Concave mirrors are used by the doctors in ophthalmoscopes, for the medical examination of ear, nose, throat, and eyes.

(iii) Concave mirrors are used in searchlights and spotlights. They are also used in the headlights of automobiles.

SUMMARY

— For all practical purposes we consider that light propagates in a straight line. This is known as rectilinear propagation of light.

— A rectangulat box having a black coating and containing a very small hole on one side and a frosted glass screen on the opposite side is known as a pinhole camera. The image formed by a simple pinhole camera is smaller than the object, it is inverted, and real.

— When light travelling in one medium strikes the boundary of another medium a part of it is sent back in the same medium, this is known as reflection.

— When a parallel beam of light is incident on a smooth surface, therefore, it is reflected in a particular direction. This is known as regular reflection. In the case when the angle of incidence is different for different rays the reflected light is scattered. This is known as irregular reflection it takes place on rough surfaces.

(i) The incident ray, the normal and the reflected ray all lie in the same plane and (ii) the angle of incidence is equals the, angle of reflection. The image of a real object formed in a plane mirror is found to be (i) of the same size as the object, (ii) it is erect, (iii) it is virtual, (iv) it is inverted laterally and (v) it is located at the same perpendicular distance behind the mirror as the object is in front.

— A real image is one that can be shown on a screen. A virtual image can not be obtained on a screen. A virtual image is formed when the rays are extended backwards behind the reflecting surface.

— A spherical mirror is considered as a section of a hollow sphere. If the inner side of the Section is polished to reflect light, it is called a concave mirror. If the outer side of the section is polished to reflect light it is called a convex mirror. The concave mirror has the ability to converge a parallel beam of light whereas the convex mirror has an ability to diverge a parallel beam of light. These mirrors are also called converging and diverging mirrors respectively.

— The terms generally used with spherical mirrors are centre of curvature, radius of curvature, pole of the mirror principal axis,

principal focus, focal plane and focal length.

— The three basic rays which are used in drawing the ray diagram are (i) a ray parallel to the principal axis, it passes or seems to pass through the principal focus after reflection, (ii) A ray passing through or appearing to pass through the principal focus. It becomes parallel to the principal axis after reflection, (iii) a ray passing through or appearing to pass through the centre of curvature, is reflected back along the same path.

— In all cases when the object is placed beyond the principal focus the images formed by the concave mirror are real and inverted. The size of the image depends upon the position of the object. If the object is placed at the centre of curvature the image formed is of the same size. When the object is shifted towards the principal focus enlargement of the image takes place and the image shifts away from the centre of curvature. When the object is shifted away from the centre of curvature the image starts diminishing and moves towards the focal point. The only case in which an enlarged, erect and virtual image is formed by a concave mirror is when the object is placed within the focal length of the mirror.

— All the images formed by a convex mirror are smaller than the object, erect, and virtual. They are all located between the pole and the principal focus of the mirror.

— The mirror formula relates the object distance p, the image distance q, and the focal length f of the mirror, it can be expressed as

$$\frac{1}{f}=\frac{1}{p}+\frac{1}{q}$$

Real object and image distances are positive. Virtual object and image distances are negative.

QUESTIONS

13.1 Write answers to the questions given below:

(i) Define reflection of light. State the laws of reflection.

(ii) Explain the formation of an image by a plane mirror.

(iii) What is meant by regular and irregular reflection of light? Describe importance of irregular reflection in daily life.

(iv) Derive mirror equation using concave mirror.

(v) What are spherical mirrors? Give some uses of spherical mirrors.

(vi) What is the difference between a real and a virtual image.

(vii) How a concave mirror is used in head lights and search lights to throw light at a long distance?

13.2 Fill in the blanks.

(i) The distance between the pole and the principal focus of a spherical mirror is the __________ of the mirror.

(ii) The image formed by a __________ mirror is always virtual, erect and smaller than the object.

(iii) An object 5 cm from a plane mirror forms an image, of the object that is the same size as the object and __________ cm from the mirror.

(iv) A ray which passes through the __________ of curvature of a mirror is __________ back along the same path after reaching the mirror.

(v) The focal plane of a spherical mirror is __________ to the __________ axis and intersects this axis at the __________ focus.

(vi) In order to solve problems involving spherical mirrors distances are measured from the __________ of the mirror, and the distances of real objects and __________ images are taken as __________.

(vii) If the object is placed in front of concave mirror between principal focus F and pole P, the image is formed __________ the mirror.

(viii) The ratio of the size of image to the size of object is called __________.

(ix) The focal length of concave mirror is taken as ________ whereas the focal length of convex mirror is taken as ________.

(x) Light is a form of ________.

13.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) The image formed in a plane mirror is ________.
(a) real (b) inverted (c) virtual and erect (d) real and inverted.

(ii) If the inner surface of a spherical mirror is reflecting it is called ________ mirror.
(a) Plane mirror (b) concave mirror (c) convex mirror.

(iii) All the rays, parallel to the principal axis, falling on a concave mirror, pass after reflection through its ________.
(a) Pole (b) Principal Focus (c) Centre of curvature.

(iv) If an object is placed at the centre of curvature of a concave mirror ________ and ________ image is formed.
(a) virtual, erect and enlarged (b) real, inverted and small (c) real, inverted and equal (d) real, inverted and enlarged.

(v) If $q = 4$ cm and $p = 2$ cm, then the magnification of the mirror is.
(a) 2, (b) 0.5, (c) 4.

13.4 Pick out, true and false from the following sentences.

(i) It is due to irregular reflection of light that sunlight reaches us before sunrise and persists for sometimes even after the sunset.

(ii) In case of a convex mirror, for all object distances, the image formed is always real.

(iii) Rays of light parallel to the principal axis after reflection from a concave mirror diverge.

(iv) If the object is placed in front of a concave mirror at principal focus F, the image is formed at infinity.

(v) In spherical mirrors, the radius of curvature is twice as large as its focal length.

PROBLEMS

13.1. An object is placed at a distance of 30 cm from a concave mirror of focal length 5 cm. If the object is 5 cm high, find the position and size of the image.

(6 cm, 1 cm)

13.2. If an object is placed at a distance of 10 cm from a spherical mirror and its virtual image is formed at a distance of 5 cm from the mirror. Find the focal length and nature of the mirror.

(10 cm, convex)

13.3. An object is situated at a distance of 20 cm from a concave mirror. Find the nature and position of the image if the focal length of the mirror is 15 cm.

(Real, 60 cm)

13.4. An object is situated at a distance of 24.0 cm from a concave mirror. The focal length of the concave mirror is 6 cm. Determine the size of the image and its distance from the mirror if the object is 12 cm high.

(4 cm, 8 cm)

13.5 The focal length of a concave mirror is 10 cm. Where should an object be placed so as to get its, real image magnified twice.

(15 cm)

13.6 The radius of curvature of a concave mirror is 40 cm. Where should an object be placed so as to get its, real image magnified four times.

(25 cm)

13.7 An object is situated at a distance of 20 cm from a convex mirror of radius of curvature 20 cm. Find the position and nature of the image.

(6.66 cm, virtual)

13.8 Focal length of a concave mirror is 10 cm. If the object is situated at a distance of (i) 60 cm, (ii) 20 cm. (iii) 5 cm from the mirror, find the distance of image in each case.

(i) 12 cm (ii) 20 cm (iii) -10 cm)

CHAPTER - 14

REFRACTION OF LIGHT AND OPTICAL INSTRUMENTS

14.1 INTRODUCTION

Why does a straight stick partly immersed at an angle in water look bent? Why do the stars appear to twinkle? Why is the sky light appears before sunrise and after the sun sets? Can you explain why it happens?

As already has been discussed in Chapter 13 that light travels in a straight line while passing through a homogenous medium. However, when a ray of light travelling through one transparent medium, enters another transparent medium, at an angle other than normal incidence, it undergoes a change in direction. The instances quoted above are due to the bending of light rays while passing from one medium to another. A variety of optical instruments such as cameras, microscope, telescopes etc, utilize lenses to make use of the phenomenon of refraction for specific purposes.

14.2 REFRACTION OF LIGHT

When a ray of light enters from one medium into another obliquely it undergoes a change not only in direction, but in velocity as well. This change of direction and velocity of light as it enters from one medium into another is known as refraction of light. The refraction of light is explained in Fig. 14.1, in which AB represents a surface separating two media. When a ray of light CO passes from one medium into another, its direction AB changed to OE instead of OD. Point O is called point of incidence; ray OE is known as refracted ray. NON′ is normal to the surface AB at the point of incidence 0. The angle between incident ray and normal ($\angle$CON) is called angle of incidence. The $\angle$EON′, the angle between refracted ray and normal is known as angle of refraction. It is deduced from experiments that when light passes from a rarer medium to a denser medium (say from air to glass);

It bends towards the normal (Fig.14.2 a) and when the light passes from denser to a rare medium, it bends away from the normal (Fig. 14.2 b)

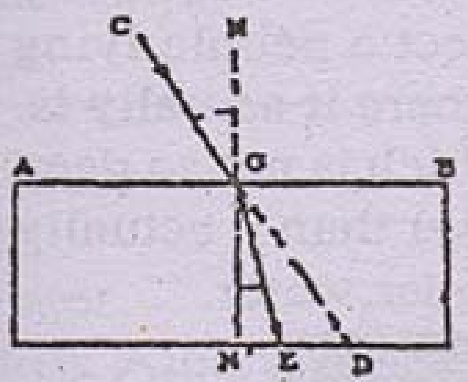


Fig. 14.1

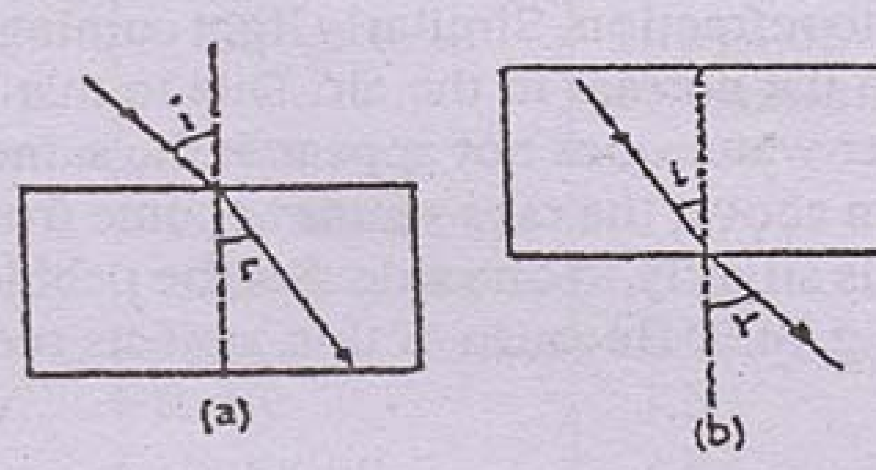


Fig.14.2

(a) When light enters from a rare medium to a denser medium it bends toward the normal (b) when light moves from denser to rare medium it bends away from the normal.

Refraction of light takes place under two laws, known as laws of refraction of light these are:

(i) The incident ray, the normal and the refracted ray at the point of incidence all lie in the same plane.

(ii) The ratio of the sine of angle of incidence (i) to the sine of angle of refraction (r) is constant for all the rays of light passing from one medium to another. This constant is called refractive index.

$$\text{Refractive Index} = \frac{\text{sine of angle of incidence}}{\text{sine of angle of refraction}}$$

$$n = \frac{\sin \angle i}{\sin \angle r}$$

This is known as Snell's Law

The constant quantity n in it is known as the refractive index of the second medium with respect to the first medium. It has no unit.

For a ray of light passing through the surface of separation of two media, the angle of incidence i is zero and sin i is also zero. As $\sin\angle i = n\sin\angle r$ therefore sin $\angle r$ is also zero because n is constant, and is not zero. This means that if a ray of light is incident perpendicularly on the surface separating the two media, the refracted ray shows no change of direction.

Refraction of light through water

When light from the air enters water it bends towards the normal due to refraction. Similarly light coming from under water bends away from the normal in the air. Due to this bending effect a pebble lying under water does not appear at the same position where it actually is. From above, the rays seems to come from a point which is not so deep and is slightly to one side. So, the pebble seems closer than it actually in Fig.14.3. Because of this, objects seem larger under water.

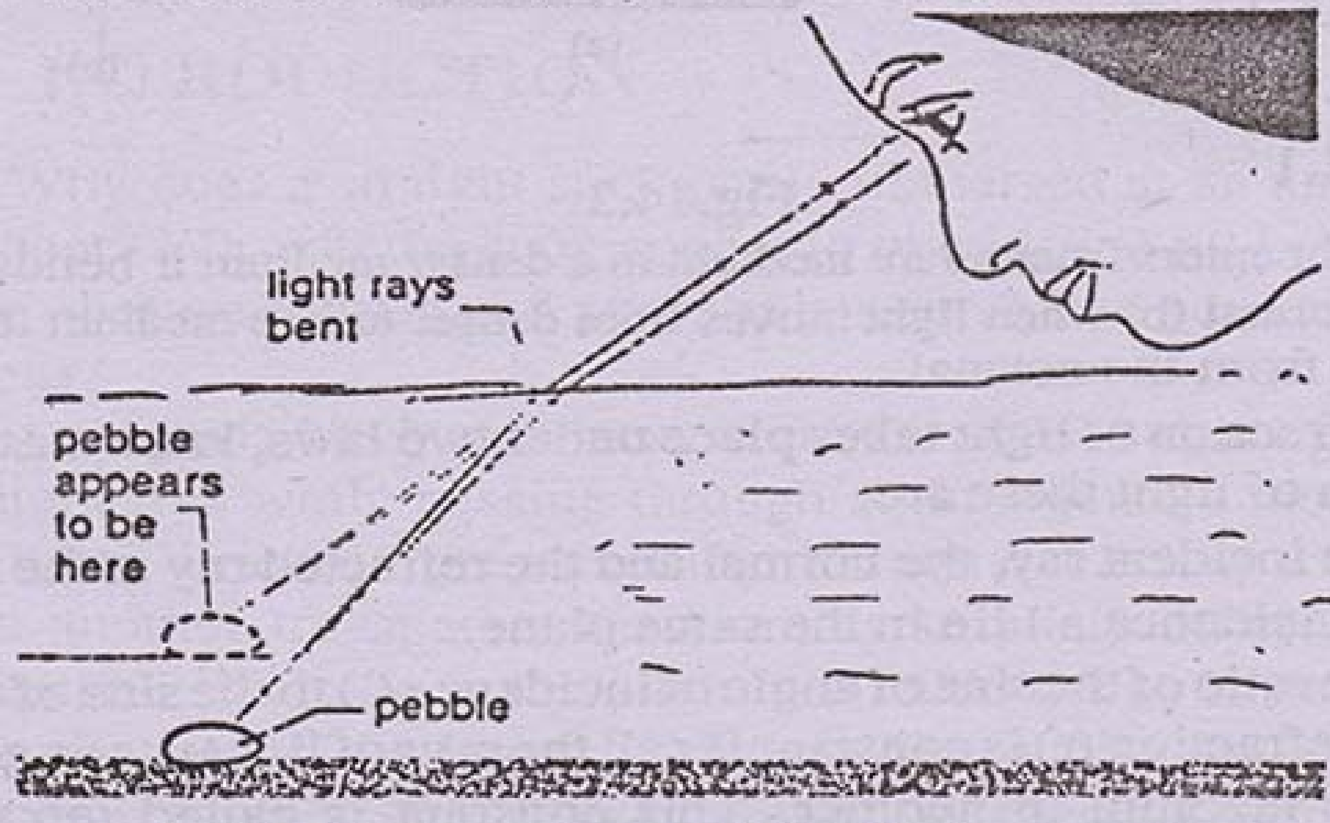


Fig 14.3 Refraction of light through water.

Refractive Index

The refractive index of a medium does not depend upon the angle of incidence. It depends upon the nature of the medium. The refractive index of a medium may also be defined as the ratio of the speed of light in vacuum to the speed of light in that medium. As the speed of light in vacuum is almost equal to the speed of light in air, therefore, we use the speed of light in air instead of speed of light in vacuum, while calculating the refractive index of a medium. For example, the velocity of light in water is $2.26\times10^{8}\,m/s$, whereas it is $3\times10^{8}\,m/s$ in air, so

$$\text{Refractive index of water} = \frac{\text{speed of light in air}}{\text{speed of light in water}}$$

$$n_w = \frac{3\times10^{8}\,m/s}{2.26\times10^{8}\,m/s} = 1.33$$

The refractive indices of some substances are given in the Table 14.1.

Table 14.1: Refractive indices of some common materials

Substance	Index of refraction	Substance	Index of refraction
Vacuum	1.0000	Glycerine	1.47
Air	1.0003,	Crown glass	1.52
Water	1.33 1	Flint glass	1.62
Ethyl Alcohol	1.36,	Ruby	1.54
Quartz	1.46 1	Diamond	2.42
Sodium Chloride	1.53	Sulphuric acid	1.43
Benzene	1.501	Carbon dioxide	1.00045

Example 14.1

The speed of light in diamond is $1.24\times10^{5}\,km/s$ and that in air is $3.0\times10^{8}\,m/s$. Calculate the index of refraction of diamond.

Solution

Speed of light in diamond $= 1.24\times10^{8}\,m/s$

Speed of light in air $= 3.0\times10^{8}\,m/s$

Index of refraction $(n_d) = ?$

$$n_d = \frac{\text{speed of light in air}}{\text{speed of light in diamond}}$$

$$= \frac{3.0\times10^{8}\,m/s}{1.24\times10^{8}\,m/s}$$

$$= \frac{3.0}{1.24} = 2.42$$

Hence, the index of refraction of diamond is 2.42

Example 14.2

Light passes from air into diamond with an angle of incidence of 45°. Calculate the angle of refraction if the index of refraction of diamond is 2.42.

Solution

The incident ray is in air, so $\angle i = 45°$ and $n_{air} = 1.00$. The refracted ray is in diamond, so $n_{dia} = 2.42$. Snell's law can be used to find the angle of refraction

$$n = \frac{\sin \angle i}{\sin \angle r}$$

or $$2.42 = \frac{\sin 45°}{\sin \angle r}$$

or $$2.42 = \frac{0.7071}{\sin \angle r}$$

or $$\sin r = 0.2922$$

$$\angle r = \sin^{-1}(0.2922)$$

Thus $\angle r = 17°$ (from the trigonometric tables)

Hence, the angle of refraction is 17°

Example 14.3

What is the index of refraction of a material if the angle of incidence in air is 60° and the angle of refraction in the material is 34°.

Solution

The index of refraction of the material can be determined from Snell's law.

From the data

$$\angle i = 60° \quad \text{and} \quad \angle r = 34$$

therefore $$\frac{n_{mat}}{n_{air}} = \frac{\sin \angle i}{\sin \angle r}$$

$$n_{mat} = \frac{\sin 60}{\sin 34} \quad (as\ n_{air} = 1.00)$$

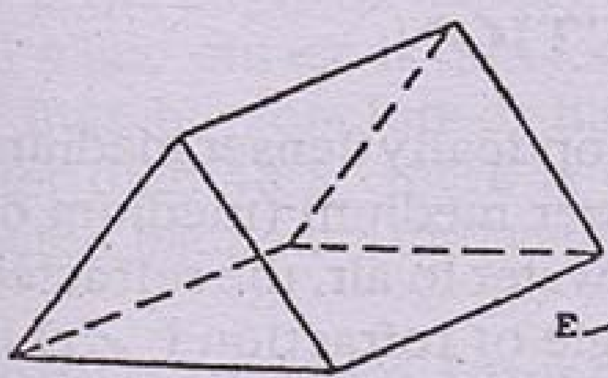
$n_{mat} = 0.8660 / 0.5592$ (from the trignometric tables)

$= 1.55$

Hence, the index of refraction of given material is 1.55.

14.3 REFRACTION OF LIGHT THROUGH A PRISM

Optical instruments such as telescopes, binoculars and periscopes use glass prisms to turn a beam of light through 90° or 180°. A prism is a transparent refracting body which is bounded by three rectangular and two triangular surfaces as shown in the Fig 14.4.



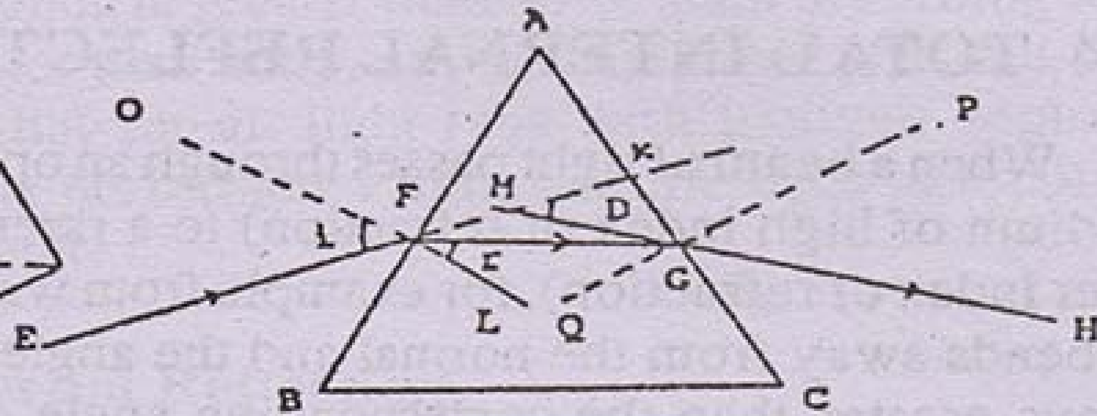


Fig. 14.4 Prism Fig. 14.5 path of a ray in a prism

The angle between the two refracting rectangular surfaces opposite to the base is called the angle of the prism ($\angle A$). Fig. 14.5 shows the path of a ray of light through a triangular prism. The ray EF strikes the face AB of the prism. On entering the prism this ray bends towards the normal OL at the point of incidence F, i.e., it bends towards the base BC of the prism. The refracted ray FG on emerging out of the prism, further bends away from the normal PQ at the point of incidence G, i.e., the emergent ray GH bends further towards the base BC of the prism. The angle EFO is the angle of incidence ($\angle i$). Angle LFG is the angle of refraction ($\angle r$). If the incident ray EF is further extended up to the point K and the emergent ray is traced back up to the point M then the angle KMG ($\angle D$) is called the *angle of deviation.* i.e., it is the angle between the extended incident ray and emergent ray. The value of the angle of deviation varies with the angle of incidence.

It may be verified experimentally that the angle of deviation is least when the incident ray and emergent ray make equal angles with The normal i.e., when the refracted ray (fig. 14.5) is parallel to the base BC of the prism. The minimum value of the angle of deviation is called the *angle at minimum deviation* and is denoted as D_m.

The index of refraction of the material of the prism can he determined by the relation,

$$n_m = \frac{\sin\frac{(A+D_m)}{2}}{\sin\frac{A}{2}}$$

where A is the angle of prism and D_m is the angle of minimum deviation.

14.4 TOTAL INTERNAL REFLECTION

When a beam of light passes through an optically denser medium (medium of high index of refraction) to a rarer medium (medium of lower index of refraction), for example from water to air, the refracted ray bends away from the normal and the angle of refraction ($\angle$r) is always greater than the corresponding angle of incidence. We also know that the rays of light are both refracted and reflected.

Consider a beam of light which travels from water $(n_w = 1.33)$ to air $(n_a = 1.0)$. The series of diagrams in Fig. 14.6 illustrate the effect of gradually increasing the angle or incidence of the light ray.

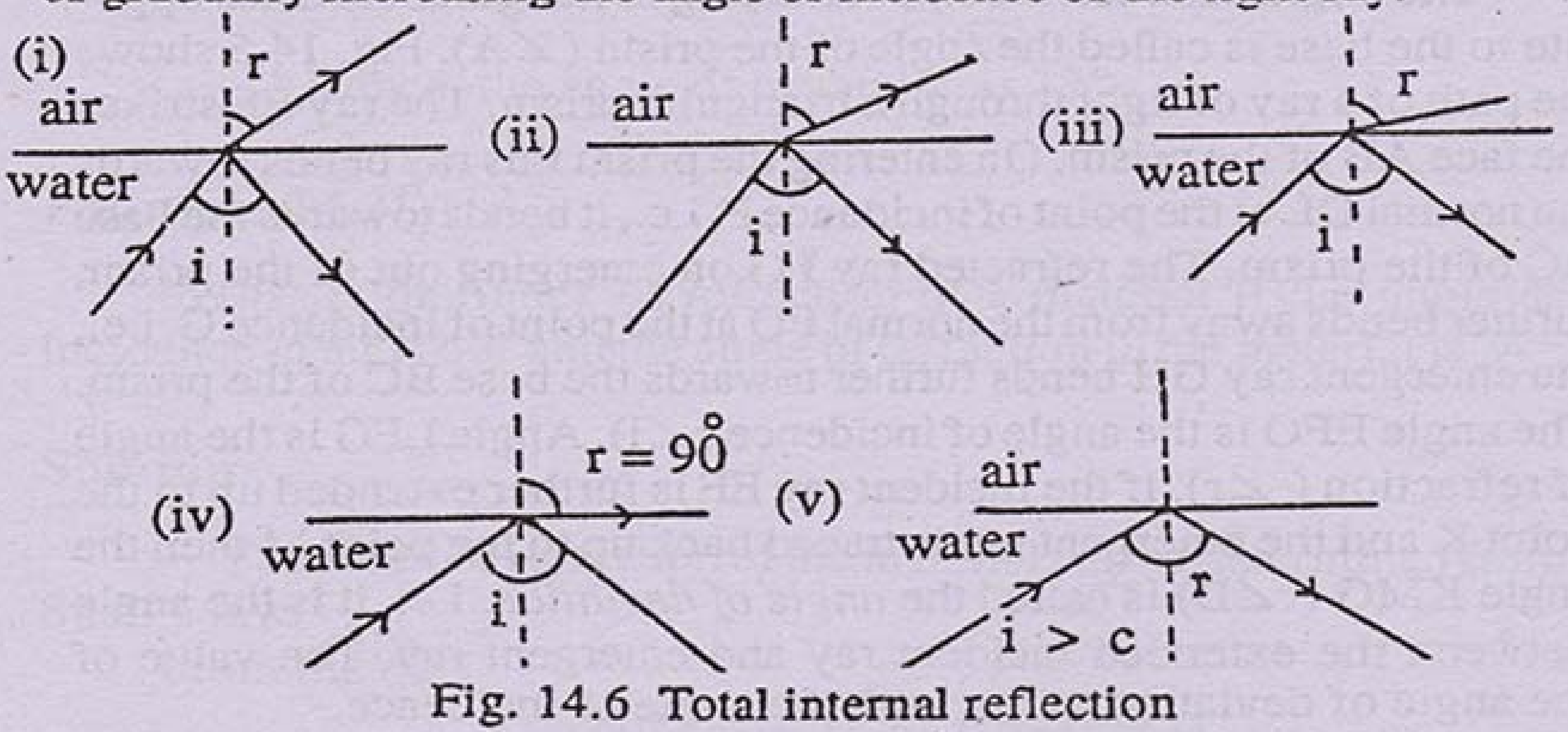


Fig. 14.6 Total internal reflection

Note that as the angle of incidence increases, the angle of refraction also increases, till for a certain value of angle of incidence, The corresponding(angle of refraction becomes 90° and he refracted ray runs along the ring surface (Fig. *14* .6). The angle of incidence for which the value of angle of refraction is 90° is called the Critical Angle (i_c). An expression for the critical angle (i_c) can be obtained from Snell's law by setting $\angle i = \angle i_c$ and $\angle r = 90°$.

When the value of the angle of incidence becomes greater than the critical angle ($\angle i_c$) there is no refracted beam but the whole ray is internally reflected back in the optically denser medium. Such reflection is called *Total internal reflection* .

For example, the critical angle for light travelling from water ($n_1 = 1.33$) to air ($n_{air} = n_2 = 1.00$) $\angle i_c = \sin^{-1}$ $(1.00/1.33) = 48.8°$. Now if the angle of incidence is greater than 48.8° then according to Snell's law $\sin\angle r$ is greater than unity, a value that is not possible. It may thus be concluded that for all light rays with incident angles greater than 48.8° there is no refracted light and the light is totally reflected back into the water.

Examples of Total Internal Reflection

1. Totally Reflecting prism.

A totally reflecting prism has one of its angles equal to 90° and each of the remaining two angles equals to 45°. If a ray of light strikes one of is faces perpendicularly, it enters the prism without any change of direction and meets the hypotenuse at an angle of 45°. As the critical angle of glass is 42° the ray striking the hypotenuse suffers total internal reflection. The reflected ray thus strikes the other face perpendicularly and comes out of the prism without any further change of direction.

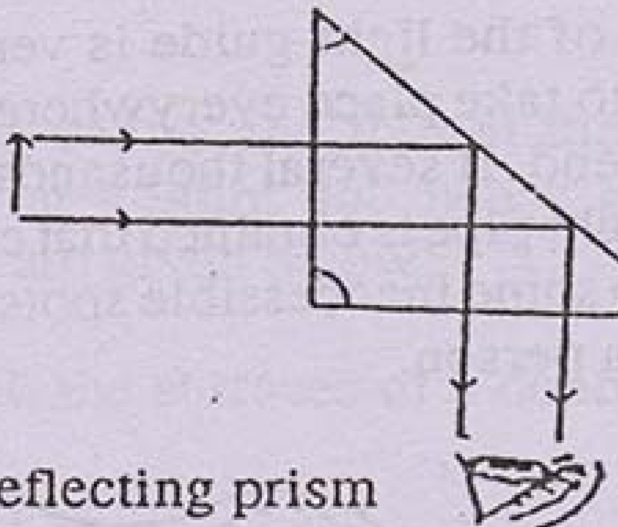


Fig. 14.7 Totally reflecting prism

It is evident from Fig. 14.7 that after passing through a totally reflecting prism the ray is deviated through an angle of 90°. This property of total reflection is used in periscopes as reflectors.

In a periscope, two 45° right angled glass prisms are used as given in the Fig. 14.8. Light entering and leaving each prism does not suffer refraction because the angle of incidence is zero.

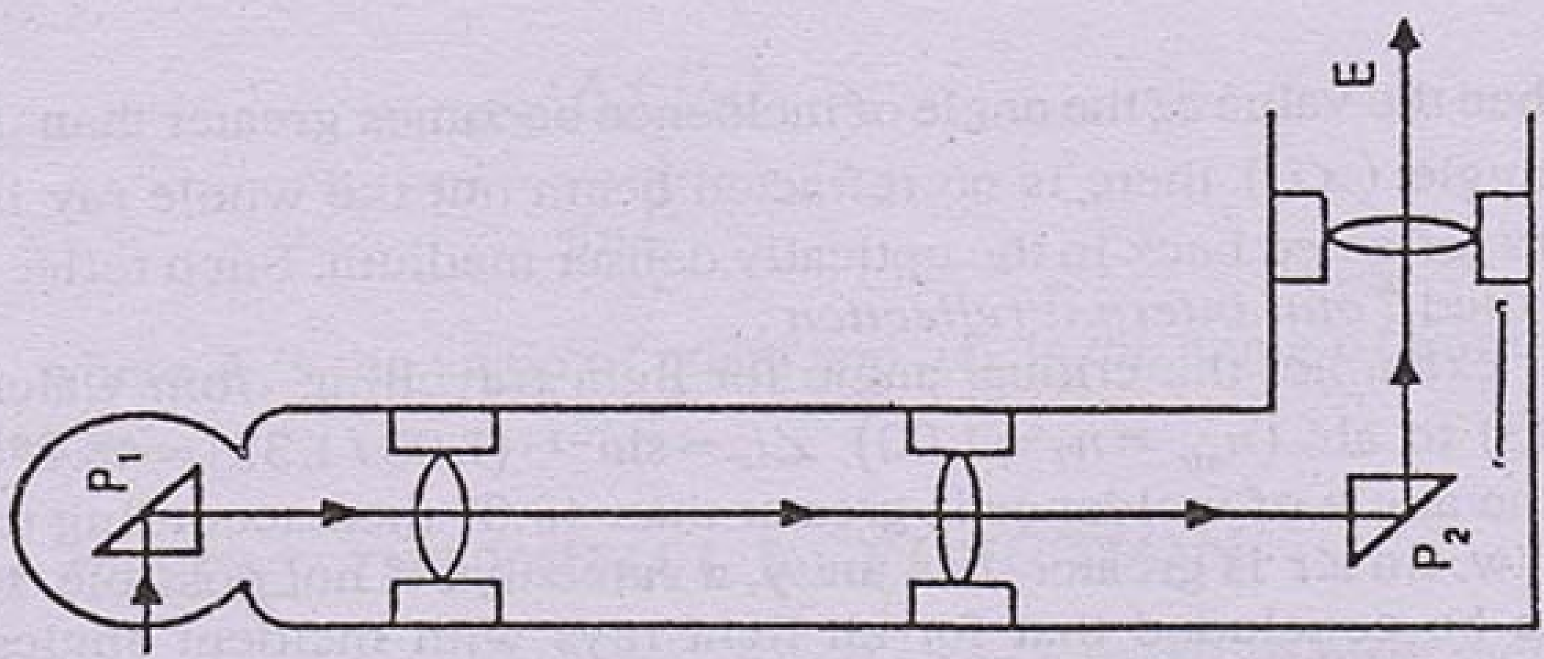


Fig. 14.8 Two prisms are used in a Periscope

Light rays striking the hypotenuse side of the prism at an angle of 45° are totally internally reflected because the value of critical angle of glass is 42° (Fig 14.8)

Totally reflecting prisms are also used in combination with a series of lenses in binoculars and projectors to change the inverted image into an erect one.

2. Optical Fibres

Another application of total internal reflection is in the exciting field or fibre optics. Light can be trapped by total internal reflection inside a very thin glass rod and travelled along a curved path (Fig.14.9). A single very thin plastic or glass fibre of about the thickness of a human hair (1/100 of a millimetre across) behaves the same way. The curvature of the light guide is very small. This allows total internal refection to take place everywhere inside the fibre. Thus light can be made to bend. If several thousands of these fibres are taped together a flexible light pipe is obtained that can be used by doctors and engineers to light up some inaccessible spots for inspection or to photograph the inside of a person.

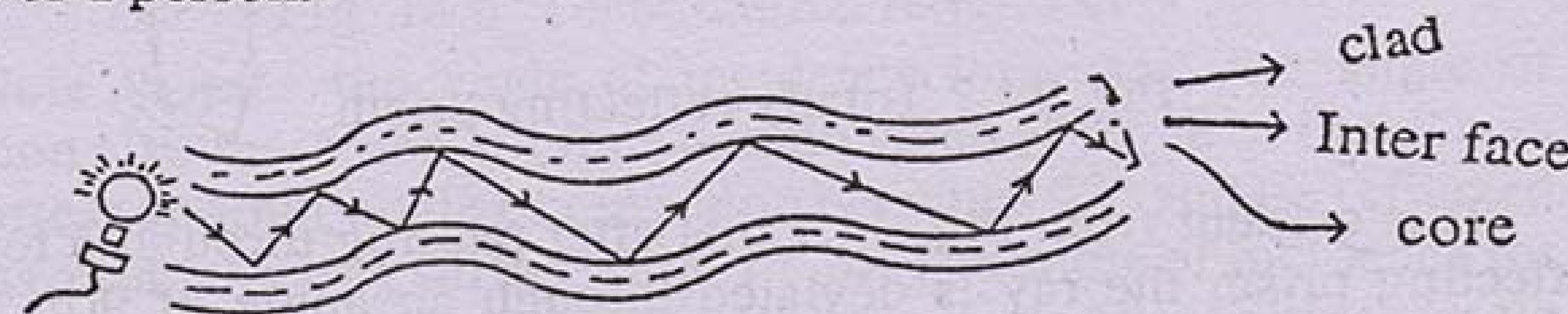


Fig. 14.9 Internal reflections in an optical fibre

In an optical fibre the light can travel with little loss because the light is totally reflected whenever it strikes the core cladding interface.

In developed countries, optical fibres are used to carry telephone signals and other modern communication systems using laser beams. A single strand of light carrying fibre can carry several thousand telephone calls at the same time without interfering with each other. Optical fibres are very light in weight, more flexible and much cheaper than copper cables. A bundle of 30 fibres are often used to form a cable. The information carrying capacity or light is thousand 1 times greater than that of electricity or radio waves.

14.5 LENSES

Millions of people wear spectacles to see clearly and to read without any difficulty. Cameras are used for photography. Magnifying glasses are used to study small objects. Projectors are used to show films on the screen. Microscopes enable us to study very tiny organism and with the help of a telescope we can clearly observe far distant objects, planets, stars and galaxies. All these optical instruments have one thing in common, that is, they all use lenses utilising the re fraction of light.

A lens is a piece of transparent material, such as glass or plastic, that refracts light in a regular way. These are bounded by one or two spherical surfaces. Two main types of spherical lenses are generally used. These are convex lens and concave lens. Convex lenses and concave lenses are of three types. They are

Convex Lenses

A convex lens is thick at the centre and thin at the edges. It converges parallel beam of light at a point and hence is called a converging lens. Their three types are given below.

(a) Double convex lens: Both the surfaces of a double convex lens are convex (Fig. 14.10 a)

(b) Plano-convex lens : One of the two surfaces of such a lens is plane and other is convex (Fig 14.10 b)

(c) Concavo-Convex lens: One of the two surfaces of a concavo-convex lens is concave and the other is convex. (Fig. 14.10 c).

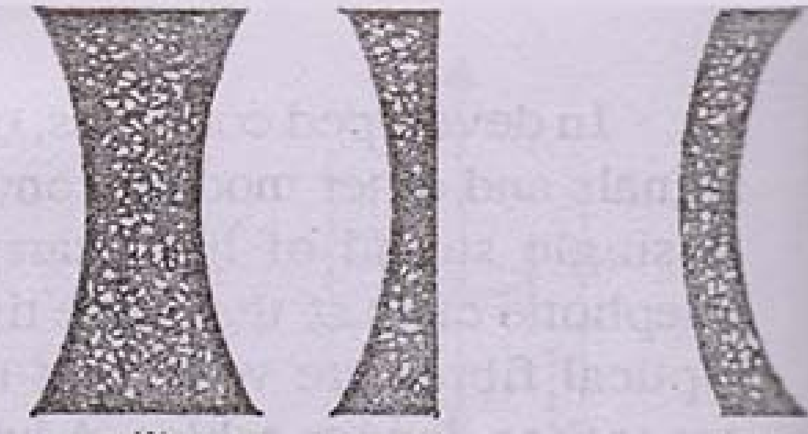


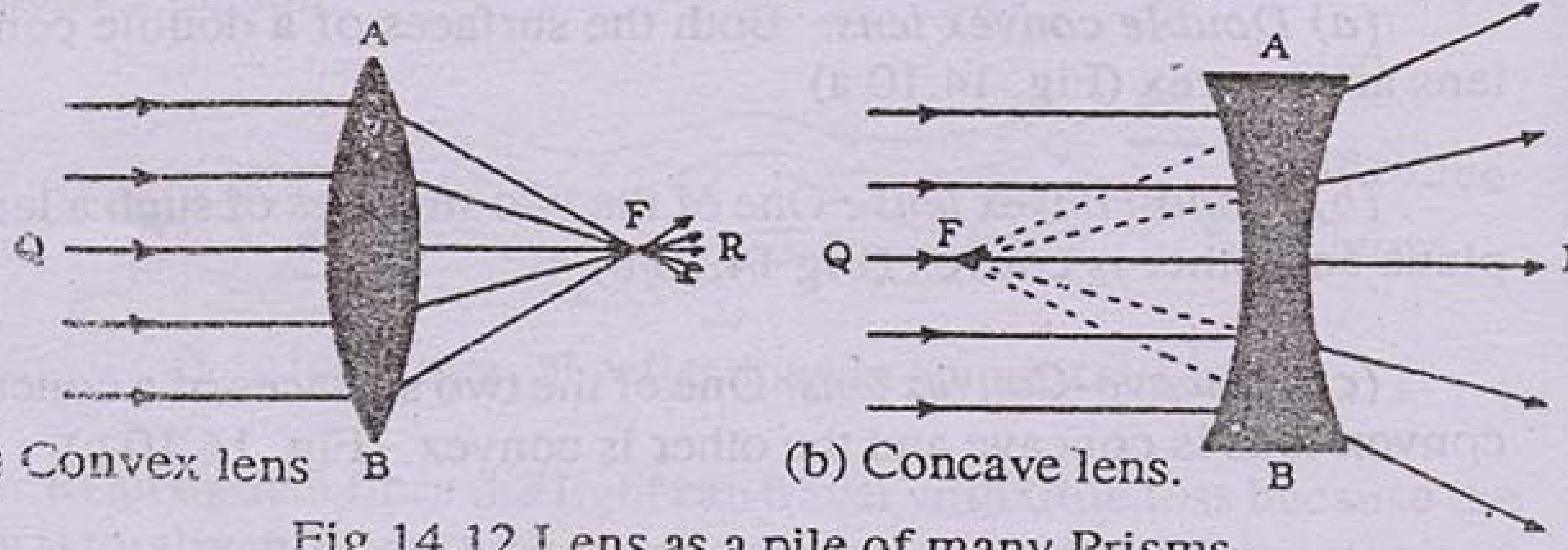
Fig. 14.10 Converging lenses   Fig 14.11 Diverging lenses.

Concave lenses

A concave lens is thinner at the centre and thicker at the edges. It diverges a parallel beam of light. The rays after refraction through a concave lenses appears to diverge from a point. The concave lens is thus called a diverging lens. There are three types of concave lens . These are, (i) Double Concave lens , *(ii)* Plano-Concave lens and *(iii)* Convex-Concave lens . These are illustrated in Fig.14.1 1.

Refraction Through Converging and Diverging Lenses: As already has been discussed a ray of light on striking the interface between the two transparent materials is either refracted towards the normal or away from it. Figs. 12. 12(a,b) show two crude lenses formed from a pile of prisms of different apex angles. Suppose an object, centred on the principal axis (A straight line joining the centres of curvature of two faces of the lens), is infinitely far from the lens so that the rays from the object are parallel to the principal axis.

When this beam of light passes through the lens each ray is refracted by a different amount at each curved surface or the prisms. The angles of incidence are greatest at the edges of the lens so bending is greatest. The bending is least at the centre as lhe opposite faces of the lens are almnst parallel. The light rays deviate in such a manner, in the case of a convex lens, that they converge to a single point on the principal axis after emerging from the lens Fig 14.12. (a).



(a) Convex lens   (b) Concave lens.

Fig 14.12 Lens as a pile of many Prisms.

This point is called the focal point F of the lens. Thus an object located infinitely far away on *the principal axis* give rise to an image at the focal point of the lens. The distance between the focal point and the lens *is* the focal length, f. This type of lens is commonly used in optical instruments. The rays emerging from the diverging lens appear to originate from a single point on the axis after passing tbrough the lens. This point is the focal point F of the diverging lens. The distance from the point F to the lens is its focal length f. This type of lens is also commonly used in optical instruments.

Graphical Construction of images using Ray diagrams: In order to understand the principle of image formation, we make use of graphical constructions. We know that an object scatters light falling on it in all directions but for the purpose of locating its image only those rays are considered which pass through the lens.

In the following we use ray diagrams to determine the location and size of the image.

We know that the light rays can pass through a lens from left to right or from right to left. Thus, the focal point of a lens can be located on either side of the lens. However, for convenience, the object is always assumed to be on the left of the lens and is oriented perpendicular to the principal axis. In order to make a ray diagram it is useful to consider two out of three rays that leave the top of the object as shown in Fig. 14.13.

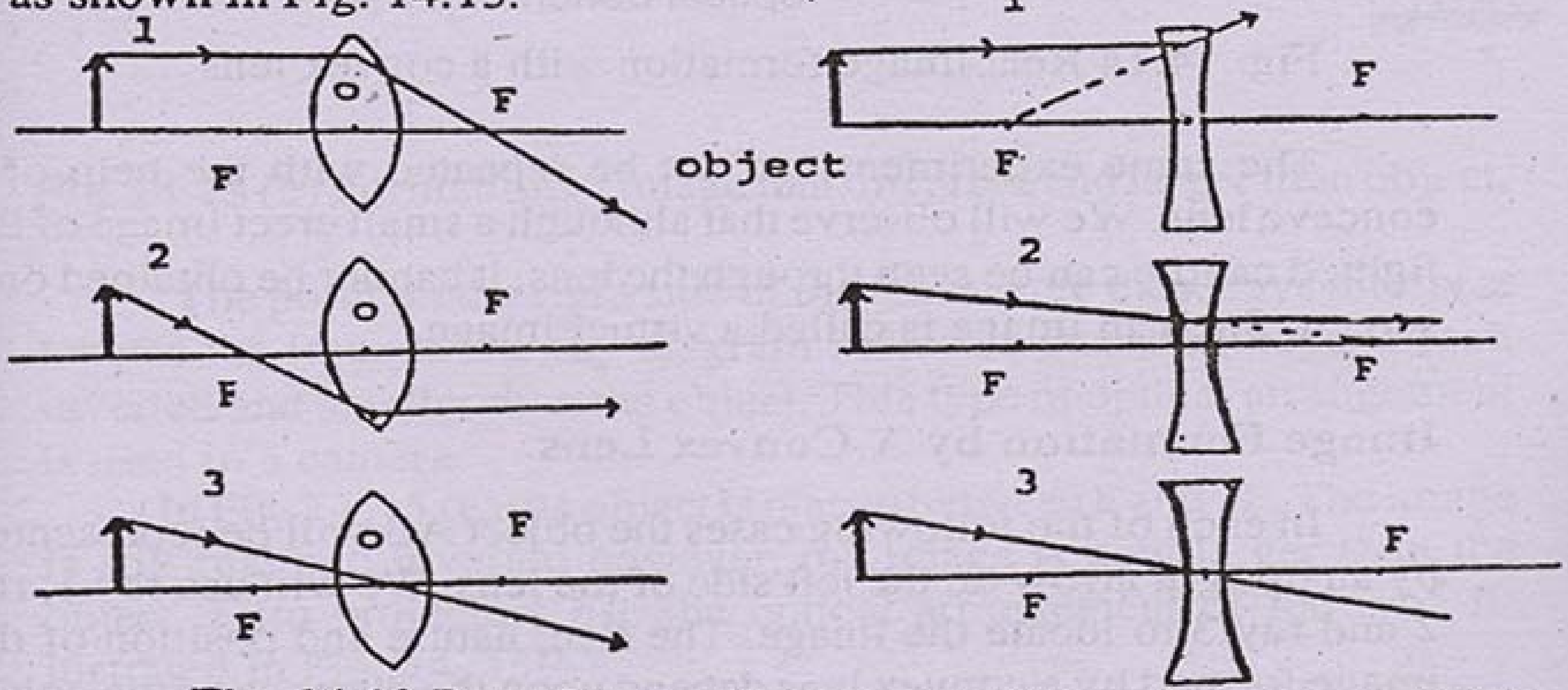


Fig. 14.13 Rays for graphical construction of images in lenses

(i) A ray parallel to the principal axis after refraction passes, or appears to pass, through the principal focus F.

(ii) A ray that passes (or appears to pass) through the principal

focus (F) at the side of the object is refracted parallel to the principal axis.

(iii) A ray that passes through the optical centre O (the geometric centre of a lens is called the optical centre), goes straight without bending because the front and back surfaces of the lens are nearly parallel at the centre and thus the lens behaves as a transparent slab.

Image Formation by Lenses

In Fig. 1 4.14, a lighted candle is placed on one side of a convex lens and a screen on the other side. The lens is slowly moved until a clear inverted image is obtained on the screen. Such an image is called a *real image.*

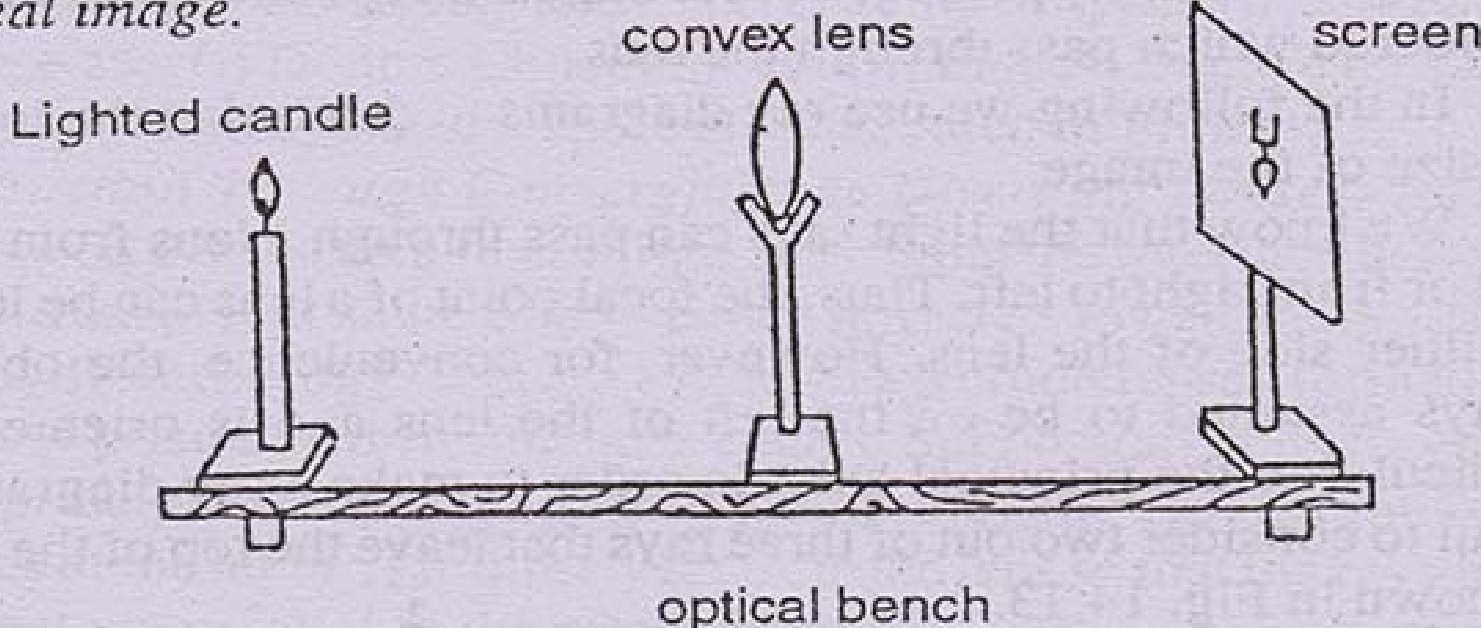


Fig. 14.14 Real image formation with a convex lens

The same experiment can not be repeated with the help of a concave lens. We will observe that although a small erect image of the lighted candle can be seen through the lens, it cannot be obtained on a screen. Such an image is called a virtual image.

Image Formation by A Convex Lens

In each of the following cases the object AB will be represented by an upright arrow on the left side of the lens. We will use ray 1, ray 2 and ray 3 to locate the image. The size, nature and position of the image formed by a convex lens depend upon the distance of the object from the lens. Fig. 14.15 (a) illustrates the formation of a real image by a converging (convex) lens. Here the object is located at a distance from the lens that is greater than twice the focal length.

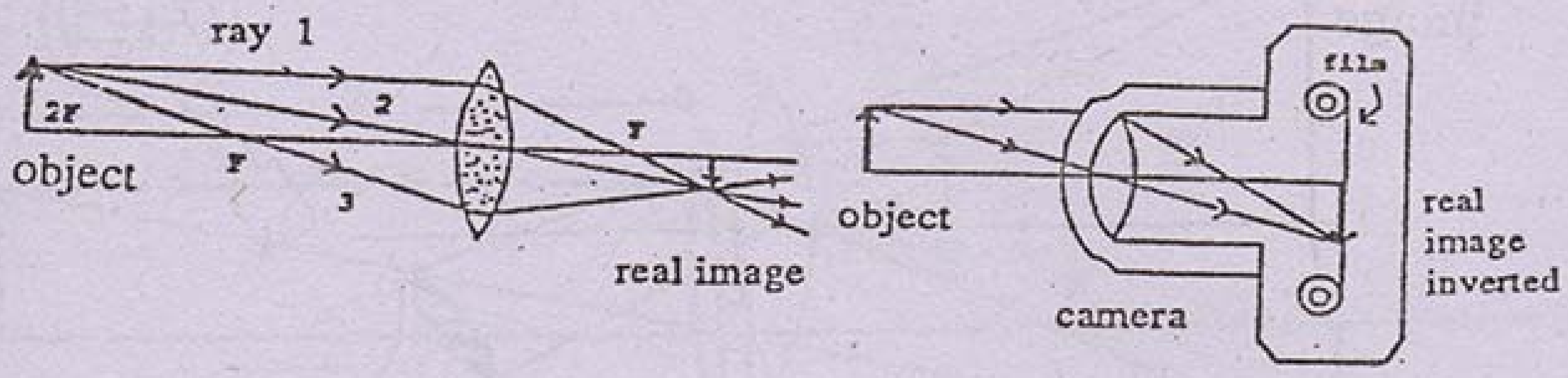


Fig. 14.15 (a) Formation of a real inverted and smaller image

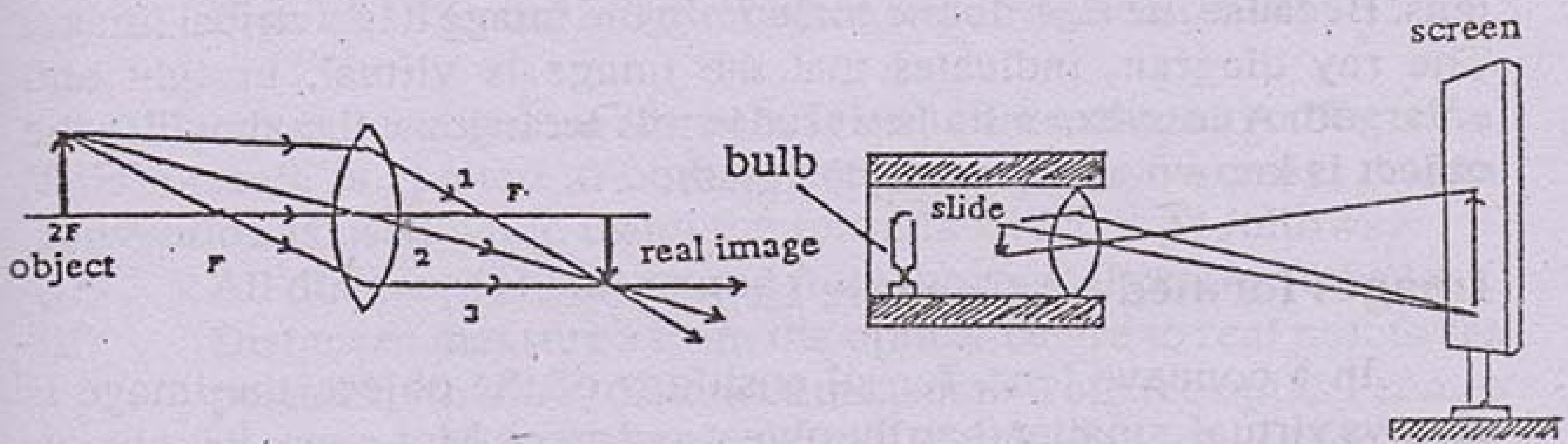


Fig. 14.15 (b) Formation of image real, inverted and larger than object.

The point on the right side of the lens where these rays intersect locates the image. The ray diagram indicates that the image is real, inverted and smaller than the object. This type of optical arrangement is used in a camera.

In Fig. 14.15 (b) the object is placed between F and 2F. The image is still real and inverted however, the image is now larger than the object. Film projectors use the optical arrangement of the type illustrated in Fig. 14. 15(b).

In Fig. 14.16 the object is placed between the focal point F and the lens: It is seen that the rays coming from the tip of the object do not intersect each other after leaving the lens.

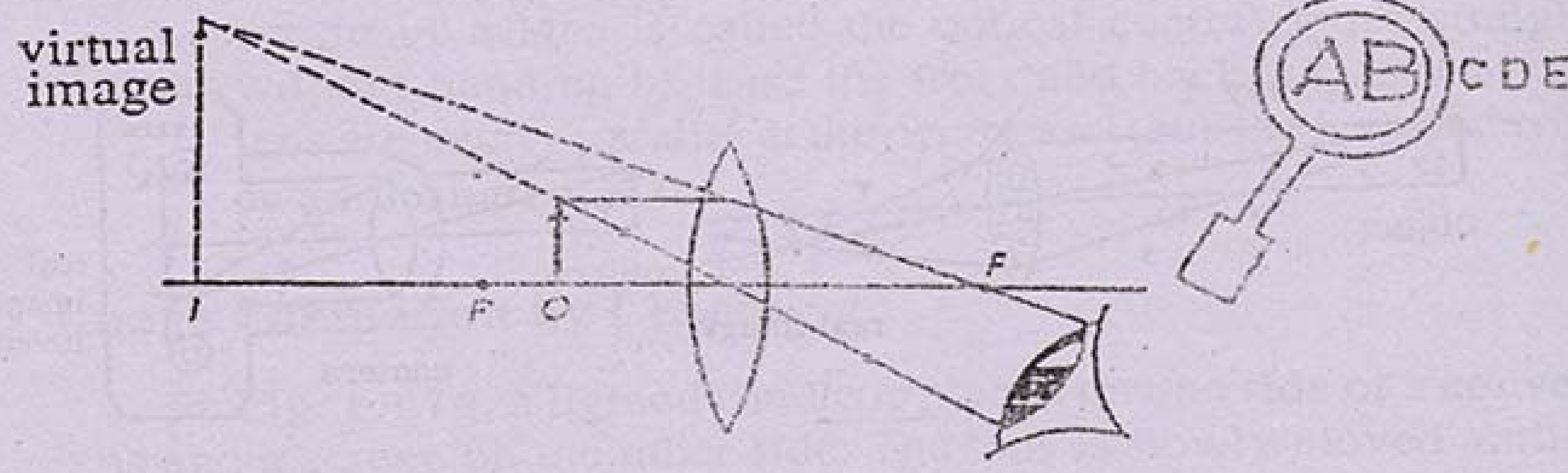


Fig. 14.16 Object between the lens and F.

They appear to come from an image behind and to the left of the lens. Because the rays do not come from the image it is a virtual image. The ray diagram indicates that the image is virtual, upright and enlarged. A convex lens when used in this arrangement to magnify the object is known as a magnifying glass.

Images formed by Concave Lens

In a concave lens, for all positions of the object, the image is always virtual, smaller than the object and erect. Moreover, it is always located between the principal focus and the optical centre as illustrated in Fig. 14.17.

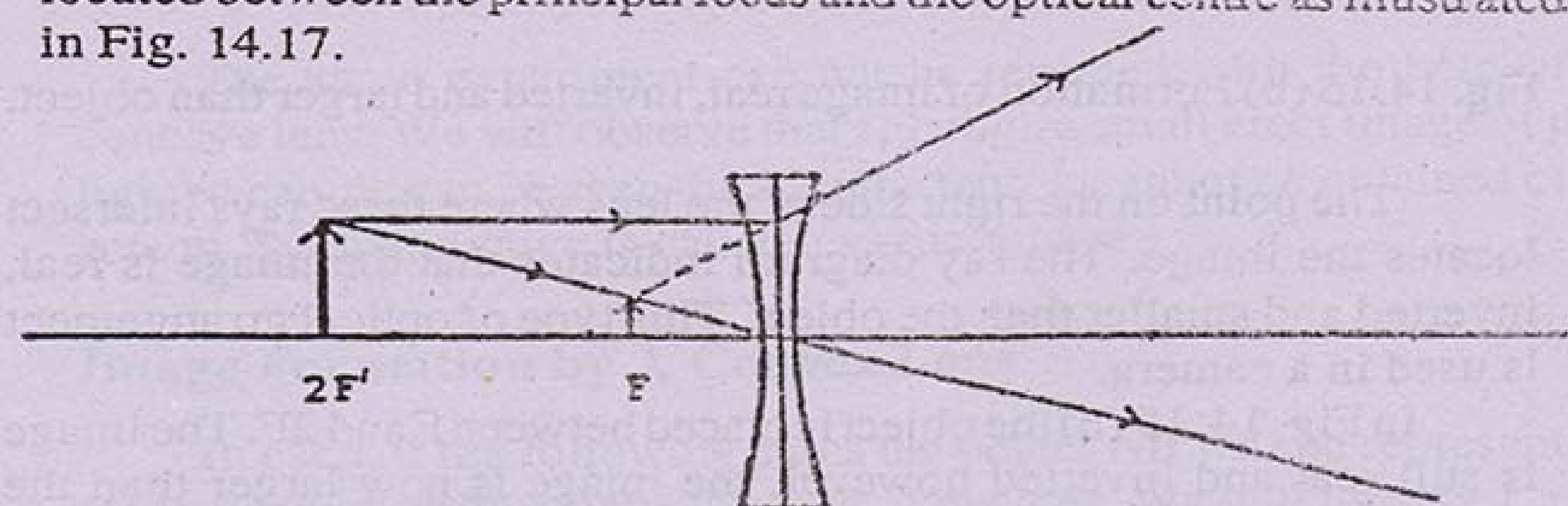


Fig. 14.17 Image formed by a concave lens

14.6 THIN LENS FORMULA

The nature and position of an image formed by a lens can also be calculated from the following formula:

$$\frac{1}{f}=\frac{1}{p}+\frac{1}{q}$$

Here p is the distance of the object from the lens, q the distance of the image from the lens, and f is the focal length of the lens (Fig.14.18).

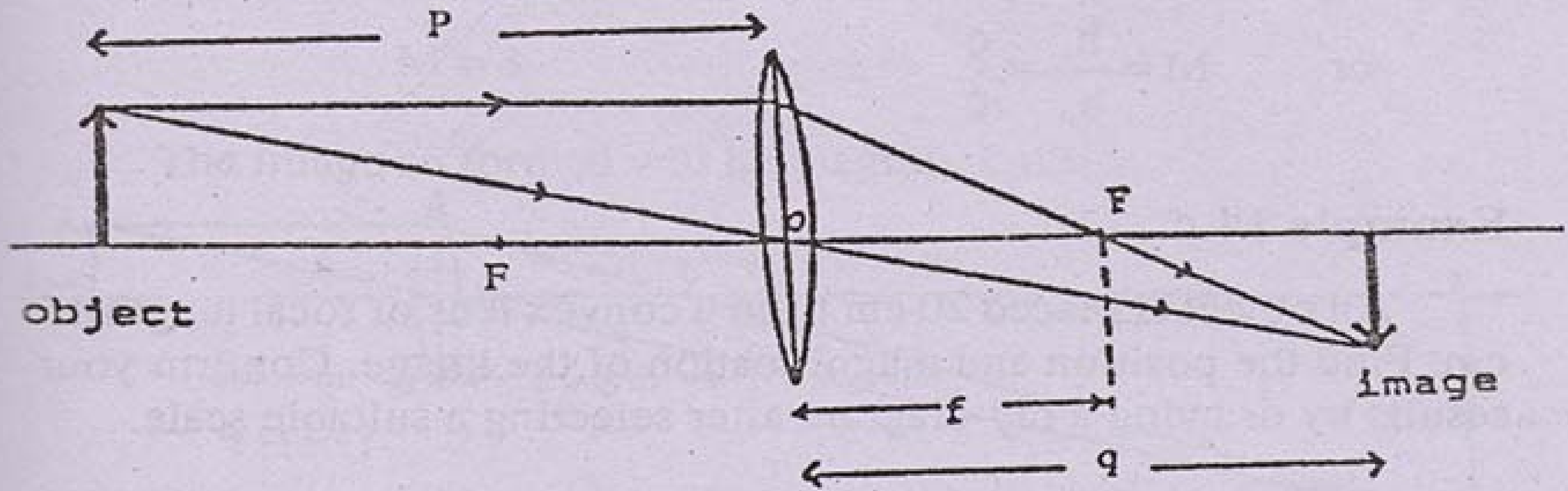


Fig. 14.18 Focal length, object and image distances

This is called thin lens formula. It is similar to the mirror formula. This formula is applied to both convex and concave lenses. A sign convention is used while using the formula which is as follows:

(i) All distances are measured from the optical centre of the lens.

(ii) Distances measured from the optical centre to real points are positive, whereas, distances measured from the optical centre to virtual points are negative. Points through which light actually passes are called real points and points through which light only appears to pass are said to be virtual points.

(iii) The focal length of a convex lens is positive and the focal length of a concave lens is negative.

According to this convention a convex lens has a real principal focus and a positive focal length, whereas a concave lens has a virtual principal focus and a negative focal length.

14.7 MAGNIFICATION

The linear or lateral magnification for lenses is determined by the following formula:

$$\text{Magnification}=\frac{\text{Height of image }(h_i)}{\text{Height of object }(h_o)}$$

$$=\frac{\text{Distance between image and lens (q)}}{\text{Distance between object and lens (p)}}$$

$$=\frac{q}{p}$$

or $$M=\frac{h_i}{h_o}=\frac{q}{p}$$

Example 14.4

An object is placed 20 cm from a convex lens of focal length 15 cm. Find the position and magnification of the image. Confirm your results by drawing a ray-diagram after selecting a suitable scale.

Solution

Object distance	p	=	20 cm
Focal length	f	=	15 cm
Image distance	q	=	?
Magnification	M	=	?

By putting the values of p, and f in the lens formula,

$$\frac{1}{f}=\frac{1}{q}+\frac{1}{p}$$

$$\frac{1}{15\text{cm}}=\frac{1}{20\text{cm}}+\frac{1}{q}$$

or $$\frac{1}{q}=\frac{1}{15\text{cm}}+\frac{1}{20\text{cm}}$$

or $$\frac{1}{q}=\frac{4-3}{60\text{cm}}$$

or $$\frac{1}{q}=\frac{1}{60\text{cm}}$$

or $$q=60\text{cm}$$

Hence, q is positive and the image formed will be real.

Applying the formula of magnification for real image i.e.,

$$M = \frac{q}{p}$$

$$= \frac{60cm}{20cm}$$

$$M = 3$$

The image so formed will be magnified thrice

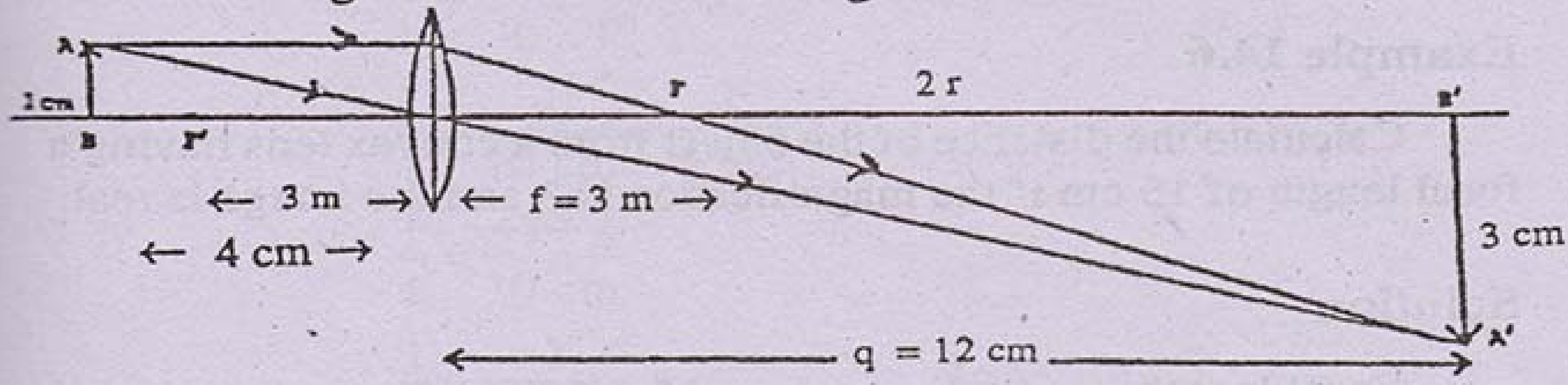


Fig. 14.19 Ray - diagram

Example 14.5

An object is placed 12 cm from a convex lens of focal length 18cm. Find the position and magnification of the image.

Solution

Object distance	P	=	12 cm
Focal length	f	=	18 cm
Image distance	q	=	?
Magnification	M	=	?

By applying the values in the lens formula

$$\frac{1}{f} = \frac{1}{p} + \frac{1}{q}$$

or $$\frac{1}{18\text{ cm}} = \frac{1}{12\text{ cm}} + \frac{1}{q}$$

or $$\frac{1}{q} = \frac{1}{18\text{ cm}} - \frac{1}{12\text{ cm}}$$

or $$\frac{1}{q} = \frac{2-3}{36\text{ cm}}$$

$$q = -36\text{ cm}$$

Since, q is negative, the image is virtual.

$$M = \frac{q}{p}$$

$$= \frac{36\ cm}{12\ cm}$$

$$= 3$$

The image is three times magnified

Example 14.6

Calculate the distance of the object from a convex lens having a focal length of 15 cm if the magnification is 3 and the image is real.

Solution

Focal length	f	=	15 cm
Magnification	M	=	3
Object distance	P	=	?

$$M = \frac{q}{p} = 3$$

or $q = 3P$

By substituting the values of f and q in the lens formula,

$$\frac{1}{f} = \frac{1}{p} + \frac{1}{q}$$

or $$\frac{1}{15cm} = \frac{1}{3p} + \frac{1}{q}$$

or $$\frac{1}{15cm} = \frac{4}{3p}$$

or $3p = 60\ cm$

or $p = 20\ cm$

Hence, the distance of object is 20 cm from the lens.

Example 14.7

Find the focal length of a convex lens if $p = 5$ cm, $q = 10$ cm and the image formed, is virtual.

Solution

Object distance $\quad p = 5\text{ cm}$

Image distance $\quad q = 10\text{ cm}$ (image is virtual, so q will be negative)

Focal length $\quad f = ?$

By substituting the values in the lens formula

$$\frac{1}{f}=\frac{-1}{q}+\frac{1}{p}$$

$$\frac{1}{f}=\frac{-1}{10\text{ cm}}+\frac{1}{5\text{ cm}}$$

$$\frac{1}{f}=\frac{-1+2}{10\text{ cm}}$$

$$\frac{1}{f}=\frac{1}{10\text{ cm}}$$

or $\quad f=10\text{ cm}$

Hence, the focal length of the convex lens is 10 cm.

Example 14.8

The focal length of a convex lens is 50 cm. An object 5 cm high is placed at a distance of (i) 25 cm (ii) 75 cm from the lens. Determine the position, nature and the height of the image.

Solution

(i)	Focal lenglh	f	=	50 cm
	Image length	h_i	=	?
	Object distance	P	=	25 cm
	Image distance	q	=	?
	Object height	h_o	=	5 cm

By substituting the values in the lens formula

$$\frac{1}{f}=\frac{1}{q}+\frac{1}{p}$$

or $$\frac{1}{50\text{ cm}}=\frac{1}{q}+\frac{1}{25\text{ cm}}$$

or $$\frac{-1}{q}=\frac{1}{25\text{ cm}}-\frac{1}{50\text{ cm}}$$

or $$\frac{-1}{q}=\frac{2-1}{50\text{ cm}}$$

or $$\frac{-1}{q}=\frac{1}{50\text{ cm}}$$

or $$q=-50\text{ cm}$$

The value of q is negative, so the image is virtual and erect.

$$-\frac{q}{p}=\frac{h_i}{h_o}\text{ (for erect image)}$$

$$-\frac{(-50)\text{ cm}}{25\text{ cm}}=\frac{h_i}{5\text{ cm}}$$

or $$25\text{ cm}\times h_1=+250\text{ cm}$$

or $$h_1=+10\text{ cm}$$

Hence the image formed is virtual, erect, 10 cm high and 50 cm towards the object.

(ii) $$f=50\text{ cm}$$
$$h_o=5\text{ cm}$$
$$p=75\text{ cm}$$
$$q=?$$
$$h_i=?$$

Therefore,

$$\frac{1}{f}=\frac{1}{q}+\frac{1}{p}$$

or $$\frac{1}{50\text{ cm}}=\frac{1}{q}+\frac{1}{75\text{ cm}}$$

or $$\frac{-1}{q}=\frac{1}{75\text{ cm}}-\frac{1}{50\text{ cm}}$$

or $$\frac{-1}{q}=\frac{2-3}{150\text{ cm}}$$

or $$\frac{+1}{q}=\frac{+1}{150\text{ cm}}$$

or $\quad q = 150 \text{ cm}$

The value of q is positive, so the image is real

or $$\frac{h_i}{h_o} = \frac{q}{p}$$

or $$\frac{h_i}{5 \text{ cm}} = \frac{150 \text{ cm}}{75 \text{ cm}}$$

or $$75 \text{ cm} \times h_i = 150 \text{ cm} \times 5 \text{ cm}$$

or $$h_i = \frac{-150 \text{ cm} \times 5 \text{ cm}}{75 \text{ cm}}$$

or $$h_i = 10 \text{ cm}$$

Hence, the image is at 150 cm, real, inverted and 10 cm high.

Example14.9

A concave lens has a focal length of 5.0 cm. If an object is placed 10.0 cm from the lens how far is the image from the lens. Confirm your result by drawing a scale ray-diagram.

Solution

Focal length $\quad f = 5$ cm (because the lens is concave)
Object distance $\quad p = 10$ cm
Image distance $\quad q = ?$

$$\frac{1}{f} = \frac{1}{p} + \frac{1}{q}$$

or $$-\frac{1}{5 \text{ cm}} = \frac{1}{10 \text{ cm}} + \frac{1}{q}$$

or $$-\frac{1}{q} = \frac{1}{10 \text{ cm}} + \frac{1}{5 \text{ cm}}$$

or $$-\frac{1}{q} = \frac{1+2}{10 \text{ cm}}$$

or $$-\frac{1}{q} = \frac{3}{10 \text{ cm}}$$

or $$-q = \frac{10 \text{ cm}}{3}$$

or $\quad q = -3.3$ cm

Since the image distance is negative, the image is virtual.

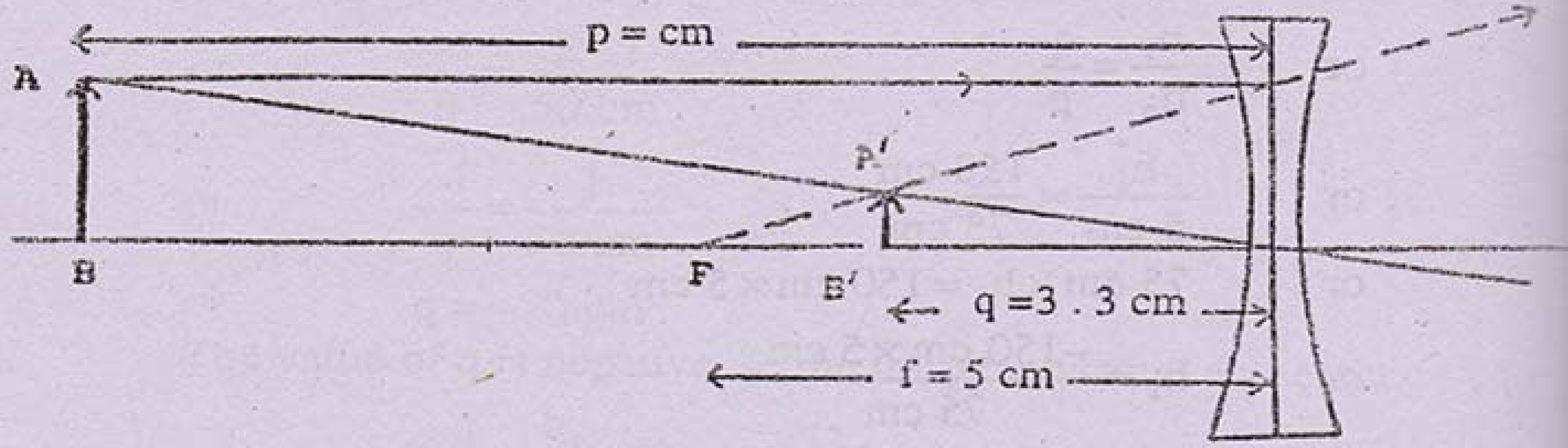


Fig. 14.20 Ray-diagram

Example 14.10

An object 10 cm high is placed 30 cm in front of a converging lens of focal length 10 cm. (a) By means of a scale ray-diagram, locate the image and find its height (b) confirm your results with the image formula.

Solution

(a)	Suppose the scale is	5 cm	=	I cm
	Object height	h_o	=	10 cm
	Object distance	p	=	30 cm
	Focal length	f	=	10 cm
	Image distance	q	=	?
	Image height	h_i	=	?

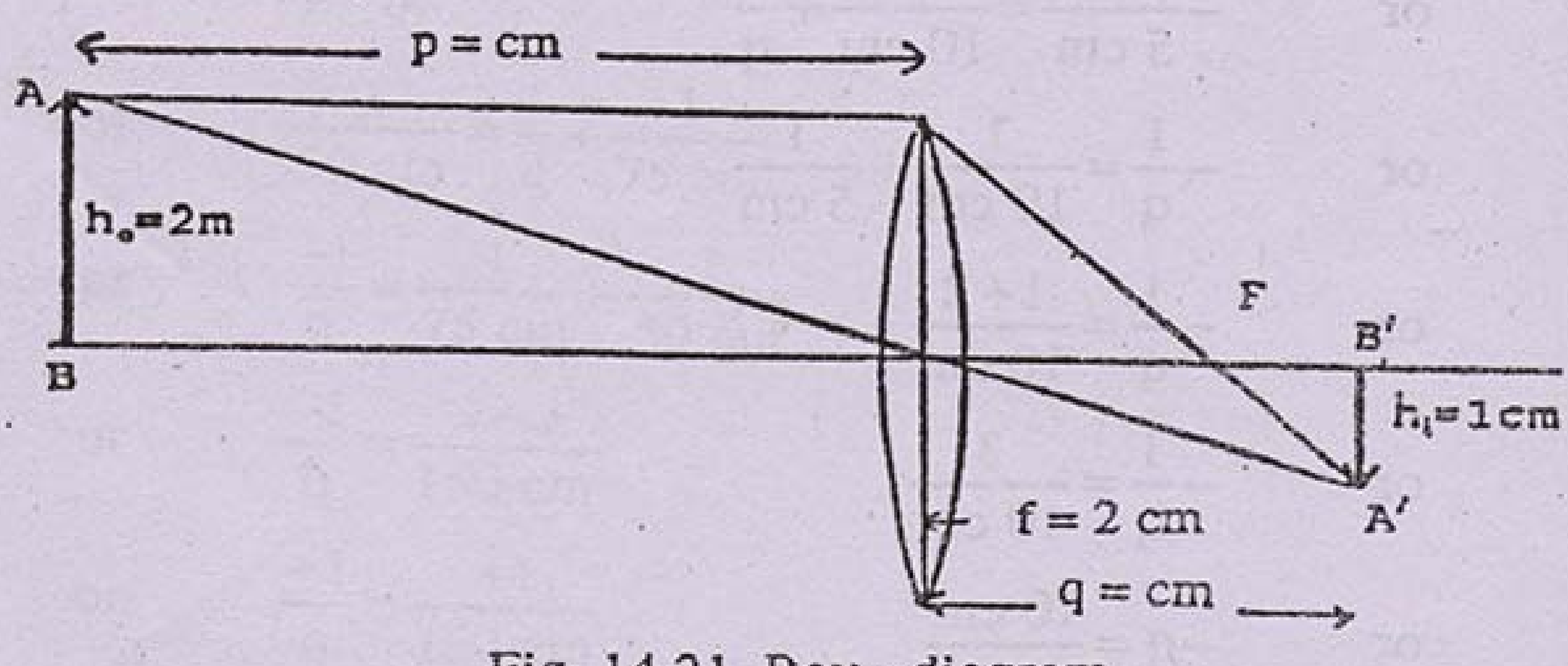


Fig. 14.21. Ray - diagram

By measurement and conversion of scale, we get,

$$q = 5 \times 3 = 15 \text{ cm}$$

$$h_i = 5 \times 1 = 5 \text{ cm}$$

(b) By using lens formula,

$$\frac{1}{f} = \frac{1}{p} + \frac{1}{q}$$

or $$\frac{1}{10 \text{ cm}} = \frac{1}{30 \text{ cm}} + \frac{1}{q}$$

or $$-\frac{1}{q} = \frac{1}{30 \text{ cm}} - \frac{1}{10 \text{ cm}}$$

or $$-\frac{1}{q} = \frac{1-3}{30 \text{ cm}}$$

or $$-\frac{1}{q} = \frac{-2}{30 \text{ cm}}$$

or $$q = 15 \text{ cm}$$

Now $$\frac{h_i}{h_o} = -\frac{q}{p}$$

or $$\frac{h_i}{10 \text{ cm}} = \frac{-15 \text{ cm}}{30 \text{ cm}}$$

or $$h_i = -5 \text{ cm}$$

14.8 OPTICAL INSTRUMENTS

All the familiar optical devices, e.g. cameras, telescopes, microscopes and eye glasses etc. function by refracting light through lenses. The lenses which are used in cameras, slide and film projectors and even the natural lens of the eye produce real images, whereas magnifying glasses are used to produce virtual images. In the case of microscopes and refracting telescopes two lenses are used. One forms a real image and the other a virtual image. In this part of the chapter we will study the construction and working of these optical instruments in some detail.

Camera

A camera is basically a light proof container blackened inside to absorb stray light. The ordinary camera has a convex lens situated at

the front, but in costly cameras a combination of lenses is used. The purpose of the lens or lenses is to produce a real and inverted image of the object (at the other end inside the camera) on a light sensitive film or plate. For sharp image of the object to be photographed, the camera is focused by moving its lens in out with the help of a mechanical mount carrying the lens. The amount of light entering the camera can be controlled with the help of a diaphragm and a shutter. A shutter of variable speed and a diaphragm with a variable aperture controls the lens of time and quantity of light entering through the lens. Normally the shutter is closed. It opens only for a fraction of a second when the button is pressed to take the photograph, (Fig. 14.22).

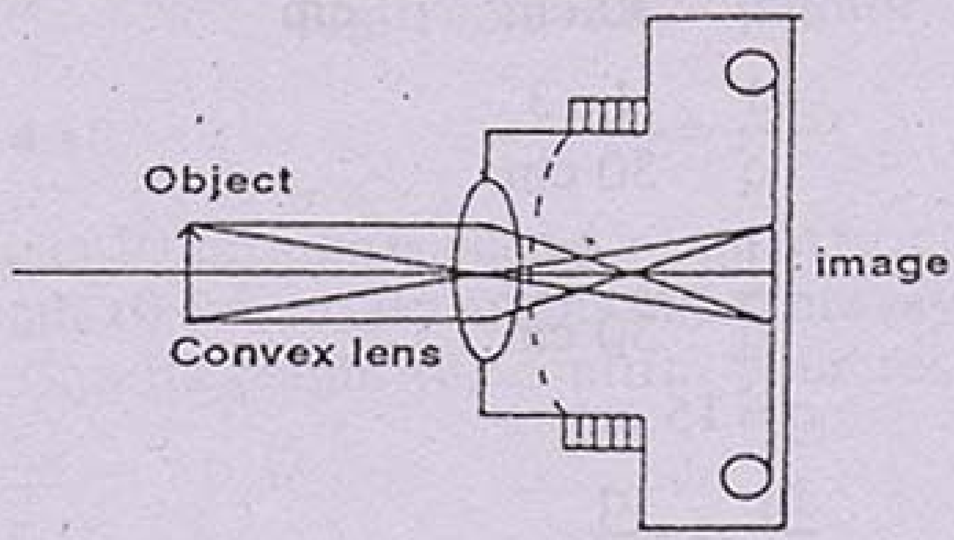


Fig. 14.22 Internal structure of camera

Human eye: Optically speaking the human eye functions in much the same manner as a camera. The eye ball is nearly spherical in structure with a diameter of about 2.5 cm. The eye ball consists of three layers namely (1) sclera (2) choroid (3) Retina enclosed in a cavity filled with a fluid called the vitreous humnur

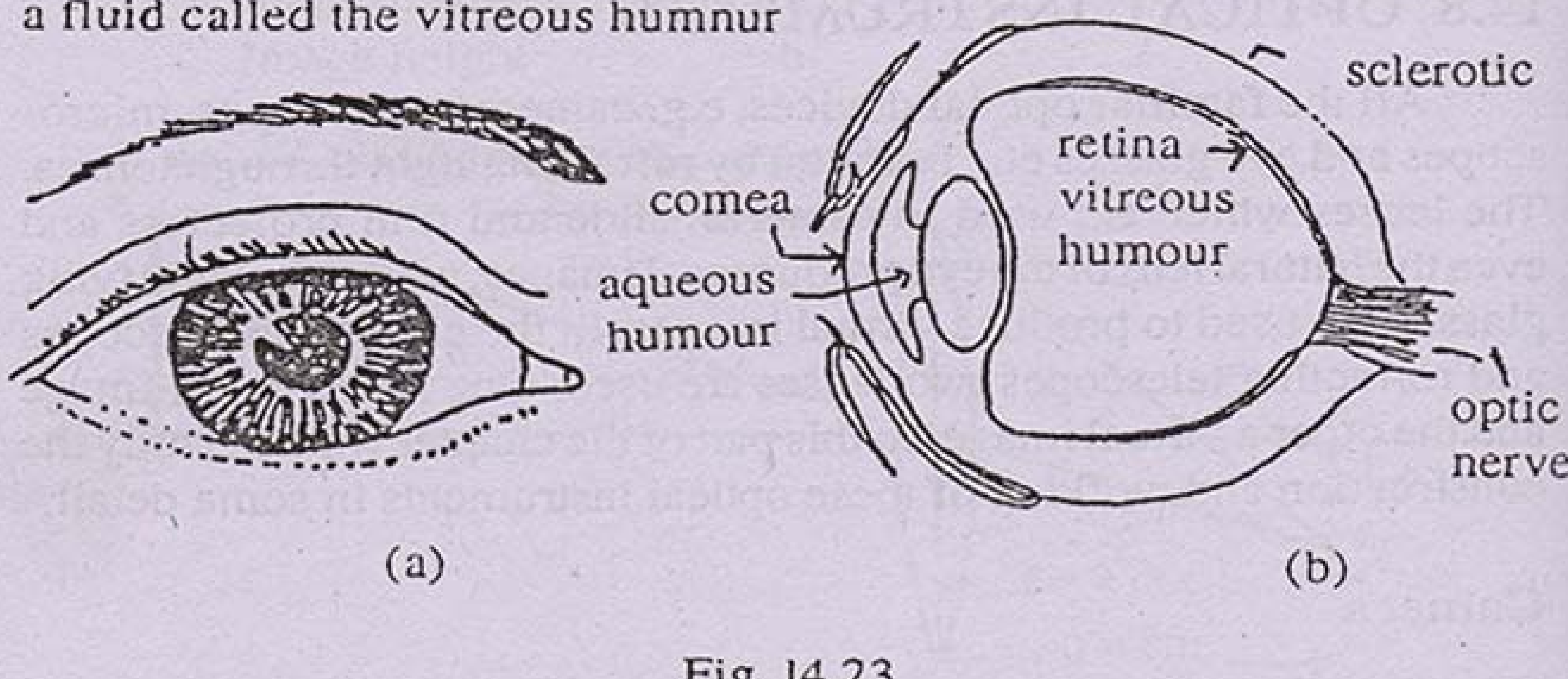


Fig. 14.23
(a) Outer view of the human eye (b) Internal structure of human eye.

The whole structure of the eye is pulled into various positions by muscles attached around the eyeball. The outer layer of the eyeball, the sclera is thick and impermeable to light rays as in a camera, but the front part of it called the cornea is transparent. The cornea allows the light to enter and the aperture of this camera is determined by the pigmented middle layer of the eye called the choroid which forms a coloured curtain called the iris which is just in front of the lens. This curtain is muscular and when it contracts, the pupil which is an opening in the centre of the iris, enlarges and more light can enter. It contracts in bright light and dilates in dim light. Its diameter varies from about 2 mm in daylight to about 6 mm in darkness. Its diameter is thus automatically adjusted to control the amount of light enteting the lens of the eye. A transparent and flexible convex lens is held just behind the iris and its focal length can be altered by the pressure of the ring-shaped ciliary muscles surrounding it (Fig 14.23 a & b). The process of changing the shape of the lens to see near by or far away objects clearly is called accommodation, (Fig. 14.24).

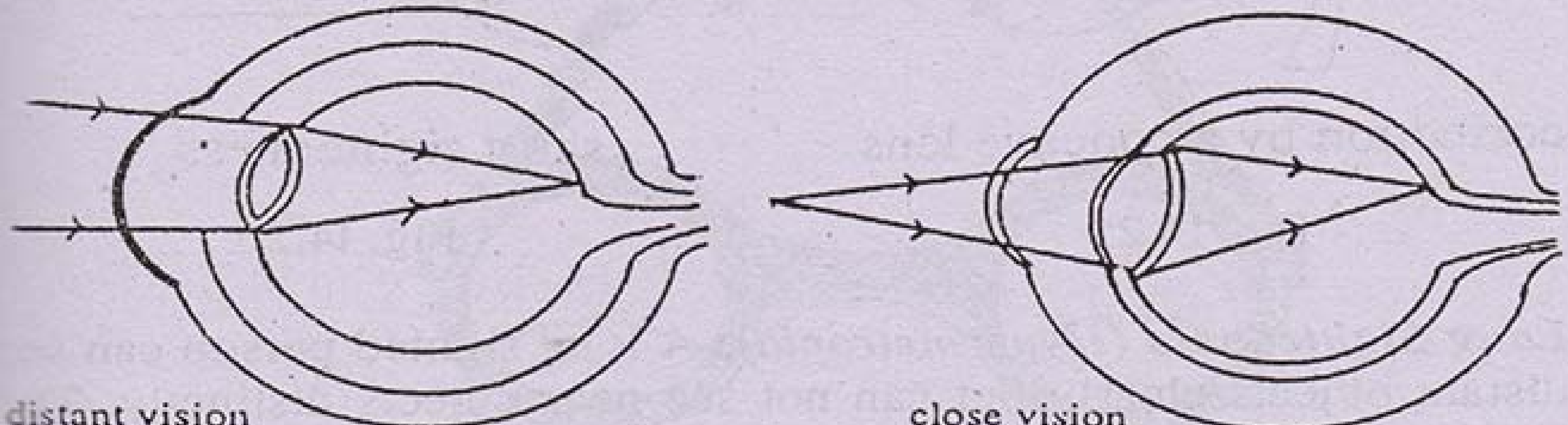


Fig. 14.24 Eye lens accomodation

The rays of light from an object are focused and converged by the lens on to the sensitive internal layer of the eyeball called the Retina which is situated at the back of the eye chamber. The retina is a nerve tissue sensitive to light. It contains over a million nerve fibres. These fibres transmit electric signals along the optic nerves to the brain. Although the image formed on the retina is inverted and reversed the brain interprets this and the images are seen the right way up.

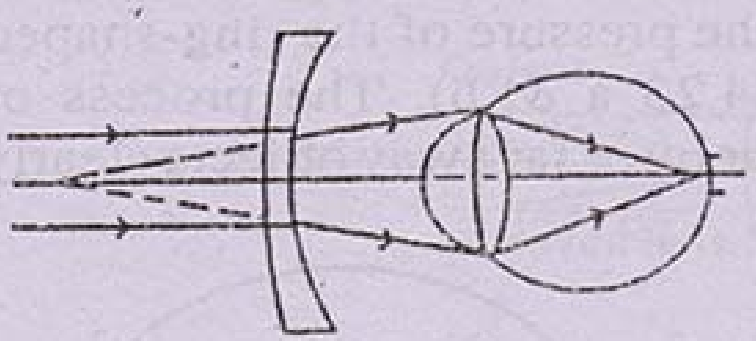
The shape of the eye is maintained by the pressure of colourless transparent fluids in the eye. The fluid between the cornea and the lens is a water-like liquid called the aqueous humour. The rest of the eye is filled with a clear jelly-like substance known as vitreous humour.

The farthest distance at which an eye can see object clearly is called the far point. The point closest to the eye at which objects can

be seen clearly is called the near point. For a normal eye, the near point is at 25 cm. This distance is called least distance of distinct vision.

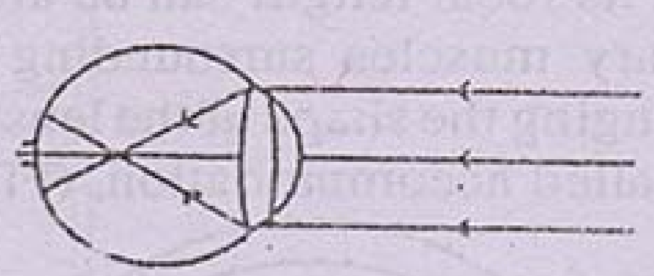
14.9 DEFECTS OF VISION

Short sightedness (Myopia): A short sighted person can see near objects clearly but distant objects are not seen clearly. The reason for this defect is either the focal length of the eye lens is too short or the eye ball is too elongated. This means that light rays from a distant object are focused in front of the retina. This defect can be corrected by wearing spectacles (or contact lenses) with concave lenses. These lenses diverge the rays of light so that the eye lens can focus the image clearly on the retina (Fig 14.25 and 14.26).



correction by a concave lens

Fig. 14.25



short sightedness

Fig. 14.26

Long Sightedness (Hypermetropia): A long sighted person can see distant objects clearly but can not see near objects distinctly. The reason for this defect is either the focal length of eye lens is too long or the eye ball is too short. This means that light rays from near objects are focussed behind the retina. This defect can be corrected by wearing spectacles (or contact lenses) with convex lenses as these lenses converge rays so that the eye lens can focus the image on the retina clearly (Fig. 14.27 and 14.28).

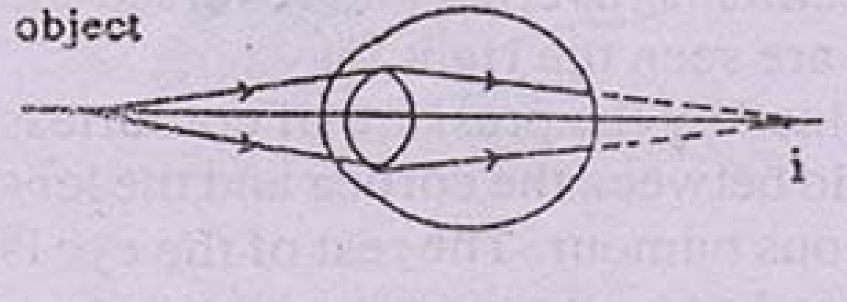


Fig. 14.27 Long sightedness

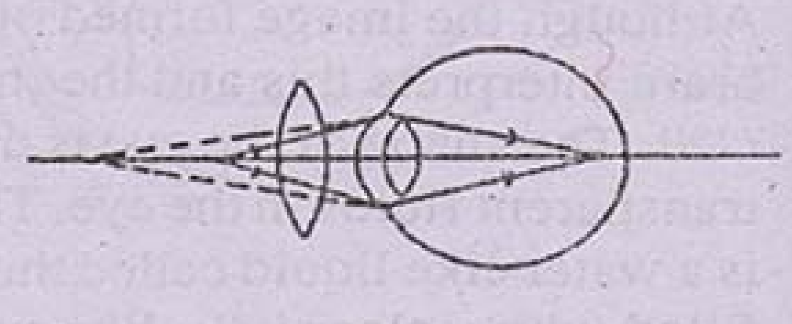


Fig. 14.28 Correction by a convex lens

Astigmatism

If the cornea, or the surface of the eye, is not perfectly spherical the eye has different focal points in different planes and an object is not focussed clearly on the retina. Astigmatism is corrected by asymmetrical lenses which have different radii of curvature in different planes.

Lack of Accommodation (Presbyopia): At old age the eye lens loses its elasticity and ability to change its shape and the ciliary muscles weaken resulting in a lack of accommodation. This kind of long sightedness is called presbyopia. This defect can be corrected by convex lenses. However, for looking at distant objects one will have to use concave lenses.

This is the reason that old people use spectacles with bi-focal lenses, i.e. convex part in the lower side to see near objects and concave part in the upper side to see the distant objects (Fig. 14.29).

Fig. 14.29 Bifocal spectacles.

Power of the Lens: Opticians and eye specialists prescribe tenses for the correction of eye defects. The unit used for the power of a lens is dioptre. The power of a lens is equal to the reciprocal of its focal length in metres.

$$P_{(\text{dioptre})} = \frac{1}{f_{(\text{metre})}}$$

For example, the power of a convex and a concave lens of 50 cm of focal length will be +2 dioptres and –2 dioptres respectively.

Simple Microscope (Magnifying Glass): A magnifying glass is simply a single biconvex (converging) lens of a short focal length (Fig 14.30). The object to be seen as magnified is placed within the focal length of

the lens. This produces, an enlarged, virtual and erect image towards the object itself (Fig. 14.31).

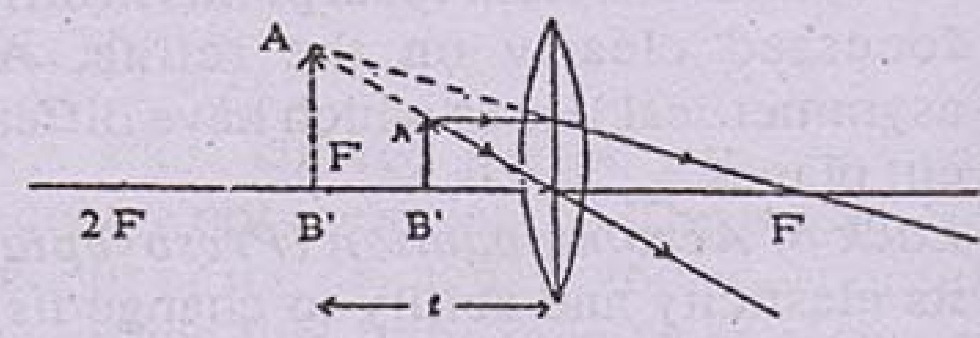


Fig. 14.30 Simple microscope Fig. 14.31 Formation of virtual image with a simple microscope.

The magnification of the magnifying glass is determined by:

$$M = 1 + \frac{d}{f}$$

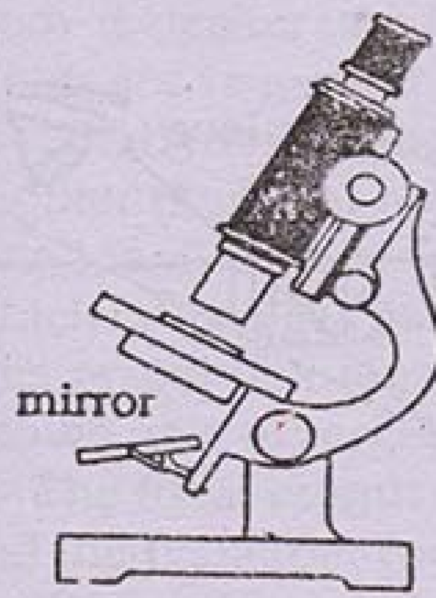
Where f is the focal length of the lens and d is the near point distance of the object which is about 25 cm for the normal eye. You can find the near point distance d for your eyes by seeing how close you can hold a page and still read it easily.

Compound Microscope: A single convex lens acts as a magnifying glass but if it is too powerful the image obtained is indistinct. If, however, two convex lenses are used a highly magnified, clear and distinct image can be obtained. This is the reason why it is called a compound microscope. The lens close to the object is called the objective and its focal length is short. The lens close to the eye is called the eye-piece. Its focal length is greater than the focal length of the objective.

The object to be viewed is placed between F and 2F of the objective lens, inverted on the stage of the microscope. The mirror at the base reflects light on the object. In most compound microscopes two or more objective lenses of different focal lengths are mounted on a rotating disc called the nose-piece. One objective lens is used at a time (Fig.14.32). The objective lens produced an inverted, enlarged and real image I_1, which acts as the object for the second lens, i.e. the eye piece. This image is focused within the focal length of the eye-piece resulting in an erect, highly magnified and virtual image I_2. This image can finally be seen by the eye. The focusing of the final image

is achieved by mounting the eye-piece in a tube that can be adjusted up and down with the help of a geared wheel (Fig. 14.32).

Each objective lens has a different power of magnificatian. Different magnification can be obtained by changing the objectives or the eye-piece or both.



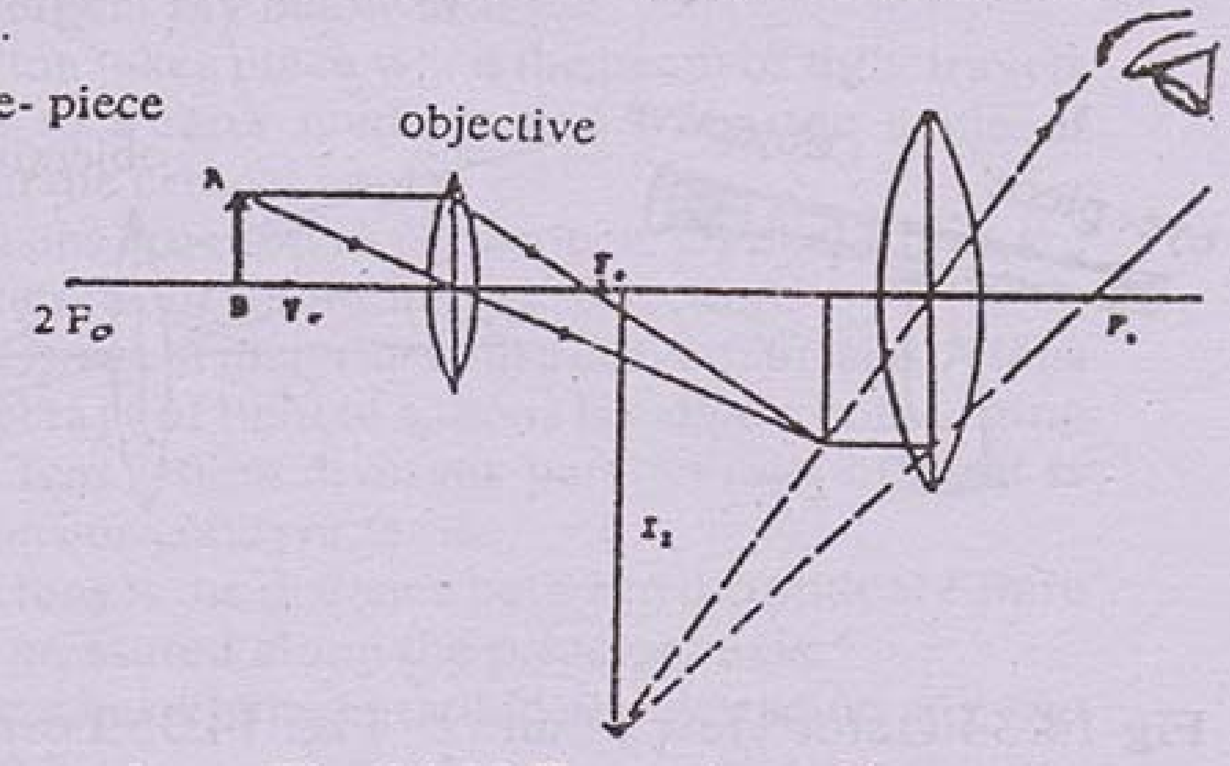


Fig.14.32 External view of a compound microscope.

Fig. 14.33 Formation of image in a compound microscope.

Refracting Telescope: An astronomical refracting Telescope consists of two convex lenses. The lens towards the object is called the objective lens. It has a long focal length. The lens nearer the eye is called the eye-piece. It has a short focal length (Fig. 14.34).

Since the stars are so distant, the rays of light coming from them will be almost parallel and are focussed to a point by the objective lens at its principal focus and form the image I_1 of the star. It is real, inverted and diminished. The eye-piece 4 is adjusted so that the image obtained from the objective is formed at the focal length of the eye-piece, which produces a virtual image I_2 of a very large magnification. The image is inverted with respect to the object but for astronomical purposes it does not matter (Fig. 14.35).

For greater magnification the objective should be of long focal length. The image formed finally is virtual so it can not be obtained on a screen behind the eye piece. However, the distance between the two lenses is made slightly greater than the sum of their focal lengths, the eye-piece no longer produces a virtual image and a real image can be obtained on a screen and we can get photographs of distant objects. No longer does an astronomer spend much time looking through a telescope with his eyes. Instead, a camera is attached to the telescope to take photographs of distant objects.

The Yerkes refracting telescope is the largest of its kind in the world. The diameter of its objective lens is about one metre. The telescope is about 18 m long and is located at William Bay, Lake Geneva, Wisconsin.

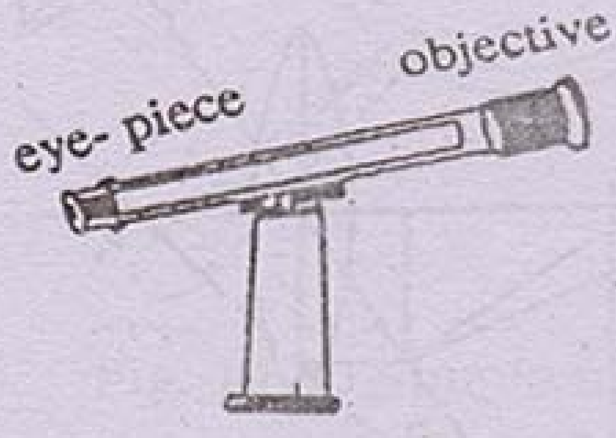


Fig 14.34 Outer view of an astronomical telescope.

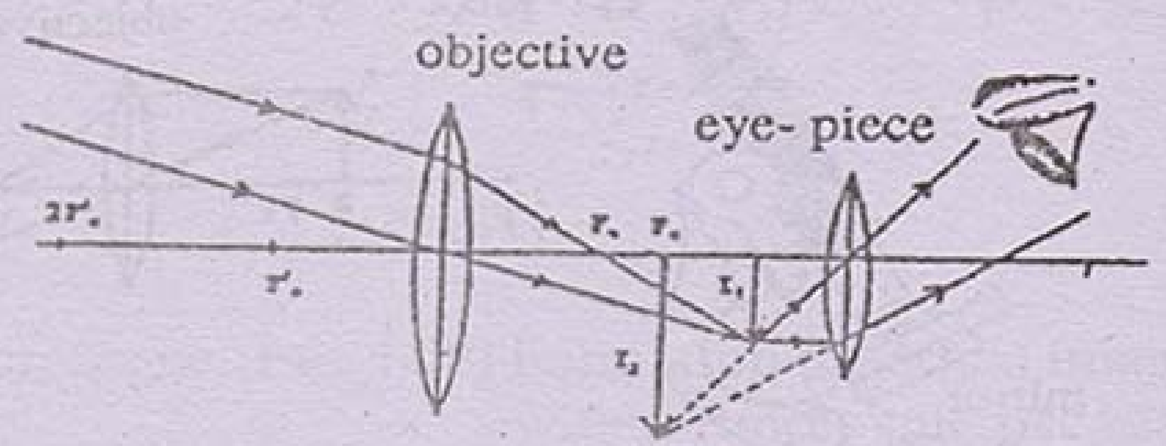


Fig. 14.35 Formation of images with an astronomical telescope.

SUMMARY

— When a beam of light travelling through one medium strikes obliquely the surface of another medium it changes its direction in the second medium. This is called refraction. When light passes from a more optically dense medium to a less optically dense medium it is refracted away from the normal drawn at the point of incidence and vice-versa.

— There are two laws of refraction. (1) The incident ray, the normal and the refracted ray all lie in the same plane. (2) The ratio of the sine of angle of incidence to the sine of the corresponding angle of refraction is constant for all rays of light passing from one medium to another.

— This above mentioned ratio is called the index of refraction or refractive index and is represented by n.

$$n = \frac{\sin \angle i}{\sin \angle r}$$

— The index of refraction can also be calculated by dividing the speed of light in vacuum (or air) by the speed of light in that medium. As the speed of light is different in different substances the indices of refraction are also different in different substances.

— When light passes through parallel-faces of a transparent medium it emerges parallel to its original direction but when the faces are non-parallel the direction of the ray is altered as in the case of a prism. The angle of deviation in a prism can be obtained by producing the incident ray forward and the emergent ray backward.

— Total internal reflection takes place when the beam of light travels from a denser medium to a rarer medium and when the angle of incidence is greater than the critical angle.

— Totally reflecting prisms are used in periscopes. Optical fibres are another practical applications of total internal reflection.

— A lens is a piece of transparent material with a curved surface. A lens which converges parallel rays of light to a focus is called a converging lens or convex lens. A lens which diverges parallel rays of light is known as a diverging lens or concave lens.

— The focal length of a lens is the distance between the optical centre and the principal focus measured along the principal axis.

— The image position can be obtained graphically by tracing any two of the following three rays.

(a) A ray parallel to the principal axis is refracted through the focus.

(b) A ray that passes through the focus is refracted parallel to the principal axis.

(c) A ray that passes through the optical centre without deviation.

— The images formed by a concave lens are always virtual, diminished and erect. The images formed by a convex lens are real and inverted when the object is located beyond the focal length of the convex lens the image is erect, virtual and magnified if the object is located within the focal length of the convex lens.

— The formula for both types of lenses and magnification is as follows:

$$\frac{1}{f}=\frac{1}{p}+\frac{1}{q} \qquad \text{and} \qquad M=\frac{h_i}{h_o}=\frac{q}{p}$$

$$\text{Power of lens} = \frac{1}{\text{focal length (m)}}$$

The unit of power is called dioptre.

— Lenses are used in optical instruments like the photographic camera, simple microscopes, compound microscopes and refracting telescopes.

— A camera has one or more convex lenses. The quantity of light striking the sensitive film is controlled by the shutter and diaphragm. A real, inverted and diminished image is formed over the film.
— The human eye has a convex lens. The thickness of the eye lens is controlled by ciliary muscles. The real, inverted and diminished image is formed on the retina which is a light sensitive nerve tissues. The retina transmits the signals to the brain through optic nerves. The brain straightens the image and they are seen the right way up. The defects of the eye are corrected with the help of the lenses.
— A magnifying glass or simple microscope is a convex lens. The object to be magnified is placed within its focal length and an enlarged, erect and virtual image is viewed through it.
— A compound microscope contains two convex lenses. The object is placed between F and 2F. The objective forms a real, inverted and enlarged image and the eye-piece further magnifies it.
— A refracting telescope consists of two convex lenses. The focal length of the objective is long and the focal length of the eye-piece is short. The objective forms a real, inverted and small image which is magnified by the eye-piece.

QUESTIONS

14.1 Write answers to the questions given below:

(i) What is meant by refraction of light? State the laws of refraction.
(ii) Explain in some detail the refraction of light through a prism.
(iii) What is a critical angle and describe total internal reflection and conditions necessary for it.
(iv) Describe the changes in the position, nature and size of the image formed by a convex lens. When the object is brought from infinity towards the lens. Illustrate your answer with diagram.
(v) Explain the functions of human eye. Name its different parts.
(vi) What are the main defects of a human eye? How are they removed?
(vii) Describe the construction and action of the following:
(a) Periscope (b) composed microscope (c) Astronomical telescope.

14.2 Fill in the blanks.

(i) When a ray of light enters obliquely from a rarer into a denser, medium then it bends __________ the normal.

(ii) The refractive index of a medium does not depend upon the angle of incidence but it does depend upon the nature of the __________.

(iii) When a ray of light enters from a denser into a rarer medium then that angle or incidence for which the corresponding angle of refraction is 90° is called __________.

(iv) Least distance of distinct vision for a normal human eye is __________.

(v) A transparent medium bounded by three rectangular surface and two triangular surfaces are called a __________.

(vi) The refractive index of the medium of a prism can be found by using the formula.

$$n = \frac{\sin ----}{\sin \frac{A}{2}}$$

(vii) A transparent medium bounded by one or two spherical surfaces is known as a __________.

(viii) A lens through which the rays after refraction converge at a point is called __________ lens.

(ix) Any ray of light passing through a particular point inside a lens does not suffer any change in direction. This point is called __________.

(x) That point where all the rays coming parallel to the principal axis, converge after passing through a lens is called __________.

(xi) The index of refraction of a substance can be calculated by dividing the speed of light in vacuum by the speed of light in __________.

(xii) When the object is at a longway away the image from a convex lens is obtained at __________.

(xiii) The image formed is always virtual, diminished and erect in a __________ lens.

(xiv) The reciprocal of the focal length of a lens is called the __________ of the lens.

(xv) The unit of power used by opticians is called a __________.

(xvi) When an eye can see clearly only distant objects, it suffers from ________.

14.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) When a ray of light enters obliquely from one medium into an other it deviates from its original path. The phenomenon is called ________.
(a) reflection (b) smaller than (c) equal to (d) unrelated to

(ii) When a ray of light enters obliquely from a rarer into a denser medium, then angle of refraction is ________ angle of incidence.
(a) greater than (b) smaller than (c) equal to (d) unrelated to

(iii) If an objcct is situated between the optical centre and the principal focus of a convex lens, then its image is formed on the same side as that of the object. This image will be ________.
(a) inverted, real and bigger in size (b) inverted, real and of the same size as that of object (c) inverted, real and smaller in size (d) erect, virtual and large in size.

(iv) For total internal reflection the angle of incidence must be ________ the critical angle.
(a) greater than (b) smaller than (c) equal to (d) half of

(v) A convex lens is ________.
(a) thinner at the centre (b) thicker at the centre (c) a diverging lens (d) plane throughout

(vi) The eye and the camera are similar because the image formed in both is ________.
(a) real, inverted and small (b) real, erect and small (c) virtual, erect and small (d) real, inverted and enlarged

(vii) In a photographic camera, a convex lens is used because it
(a) produces a real and small image (b) gives virtual image (c) produces an enlarged image (d) forms an image equal in size to the object

(viii) The objective of a refracting telescope is a ________.
(a) double convex lens (b) convex mirror (c) plane mirror (d) concave mirror.

(ix) A single diverging lens is used in ________.
(a) a magnifying glass (b) a camera

(c) the objective lens in a telescope (d) spectacles for the correction of short-sightedness.

(x) The pupil of the eye controls ________.

(a) the focal length of the eye (b) the range of accommodation of eye (c) the distance of distinct vision (d) the amount of light reaching the eye

14.4 Pick out true and false from the following sentences.

(i) If a ray of light enters normally from one medium into the other then the angle of refraction is much larger than angle of incidence.

(ii) One angle of the totally reflecting prism is 90° while the other two angles are 45° each.

(iii) The principal focus of a concave lens is always virtual.

(iv) A virtual image is that which can be obtained on a screen.

(v) The reciprocal of the focal length of a lens is called its power.

(vi) The unit of the power of the lens is diopte.

(vii) Centre of curvature is the centre of that sphere of which the lens is a part.

(viii) The distance between the centre of curvature and the optical centre is called focal length.

(ix) The image formed by a concave lens is always real.

(x) The principal focus of a convex lens is always real.

(xi) Total internal reflection always takes place when the angle of incidence is much less than the critical angle.

(xii) A simple microscope is a bi-concave lens.

(xiii) The image in the human eye is formed at the iris.

(xiv) A compound microscope contains a convex and a concave lens.

PROBLEMS

14.1 Calculate the speed of light in glycerine if its index of refraction is 1.47.

(2.04×10^8)

14.2 The speed of light in water is $2.25 \times 10^5\, km/s$. What is the index of refraction of water?

(1.33)

14.3 Light travels from air into water whose index of refraction is 1.33. If the angle of incidence is 40°. What is the angle of refraction?

(28.90°)

14.4 The focal length of a convex lens is 10 cm. Where should an image be placed to get (a) a real image (b) a vertual image twice the size of the object?

[(a) 15 cm (b) 5 cm]

14.5 Find the focal length of a convex lens if (a) $p = 5$ cm, $q = 10$ cm and the image is virtual, (b) $p = 30$ cm and the image is real.

[(a) 10 cm (b) 7.5 cm]

14.6 The focal length of a convex lens is one metre and an object is placed at a distance of 2 m before it. Determine the position, nature and magnification of the image.

(2m, real, same size)

14.7 The distance between an object and a screen is 49 cm. A convex lens is placed between the object and the screen so as to get an image on the screen magnified six times. calculate the focal length of the lens.

(6 cm)

14.8 The distance between an object and a convex lens is 18 cm. The focal length of the lens is 6 cm. Determine the nature, position and magnification by using the lens formula.

(– 4.5 cm, virtual, 0.25 times)

CHAPTER - 15

NATURE OF LIGHT AND ELECTROMAGNETIC SPECTRUM

15.1 INTRODUCTION

Light is a form of energy. Life on the earth without it would not have existed and that it enables us to see objects. We also know that it can travel through empty space and moves faster than anything else we know about. In this chapter we shall study different theories about the nature of light.

15.2 NEWTON'S CORPUSCULAR THEORY OF LIGHT

Newton in the middle of seventeenth century proposed that light consist of minute particles, called corpuscles which on emission from the source of light travel along straight line with great speed. When these particles enter the eyes, they create the sensation of sight on interacting with the retina of the eye. This is known as the Corpuscular *Theory of Light.* The theory explains the formation of shadows and propagation along straight paths in a straight-forward manner. It also explains the phenomenon of reflection by arguing that reflection of light particles from a surface takes place in the same manner as the reflection of a rubber ball from a hard surface.

Phenomenon of refraction, however, presented difficulties. Though Newton did explain the refraction phenomenon, he had to assume that velocity of light in denser media like water will be greater than in air. In 1850 A.D French Physicist Foucault proved that Newton's result, that velocity of light in water should be greater than air, is wrong. Newton's corpuscular theory was thus abondoned.

15.3 WAVE THEORY OF LIGHT

Huygen,a contemporary of Newton, proposed an alternate theory known as wave theory of light. Newton's towering personality did not allow Huygen's theory to gain acceptance for over one hundred years.

But in the later half of 19th century Huygen's theory got wide acceptance.

Christian Huygen by contrast, believed that light travelled as waves like ripples on water pond. It was known during those days that a medium was essential for the propagation of waves, therefore, it was assumed that all space was filled with a hypothetical medium called the *Ether* which enable the propagation of light waves through space.

Waves in different media propagate with different speeds and different wave lengths bend differently. It was argued by Huygen, that the waves of light were so tiny that there was no visible distortion in typical mirrors and lenses and in the formation of shadows. In this manner the wave theory of light explained the rectilinear propagation of light, formation or shadows, reflection, refraction and a few other properties of light. Later, in 1801 Thomas Young performed an experiment, which supported the wave theory of light by showing interference bands of light. Bright bands were formed when one set of waves joined another set in such a way that crests and troughs were reinforced. Dark bands were formed when these crests and troughs cancelled each other. This property of light could be explained only on the basis of the wave nature of light.

During the later half of the nineteenth century, Maxwell theoretically showed the existence of electromagnetic waves and strengthened the belief in the wave theory of light by describing the light waves as electromagnetic in nature. Experimental evidence provided by Hertz about the existence of electromagnetic waves supported the electromagnetic nature of light waves. But if light is waves, 19th century scientists wondered, how can light travel through a vacuum? Therefore, the assumption about the existence of the invisible ether as a medium for the propagation of light remained in vogue till the begining of the 20th century. Morley and Michelson then showed experimentally that ether did not exist.

15.4 QUANTUM THEORY AND DUAL, NATURE OF LIGHT

In 1905, Max Plank while studying heat radiated from a hot body, came to the conclusion that radiation was emitted in the form of tiny packets of energy. This assumption was then used by Einstein to explain the experimental observation that when the light is shone onto a metal surface electrons are emitted from it. Einstein, concluded that the physical nature of light was not that of a wave but of little packets

of light energy. He called these packets of energy "photons", thus again revived the idea of Newton's light corpuscles. But the photons of Einstein are not material particles such as balls, they can behave as waves also. This peculiar nature of light is now called the dual nature of light. It is never both at the same time. Sometimes it behaves as a particle and sometimes as a wave depending on how you look at it. The success of the photon idea explains the existing phenomena of light through empty space.

15.5 DISPERSION

The fact that sunlight consists of different colours was first investigated by Newton with the help of the following experiment.

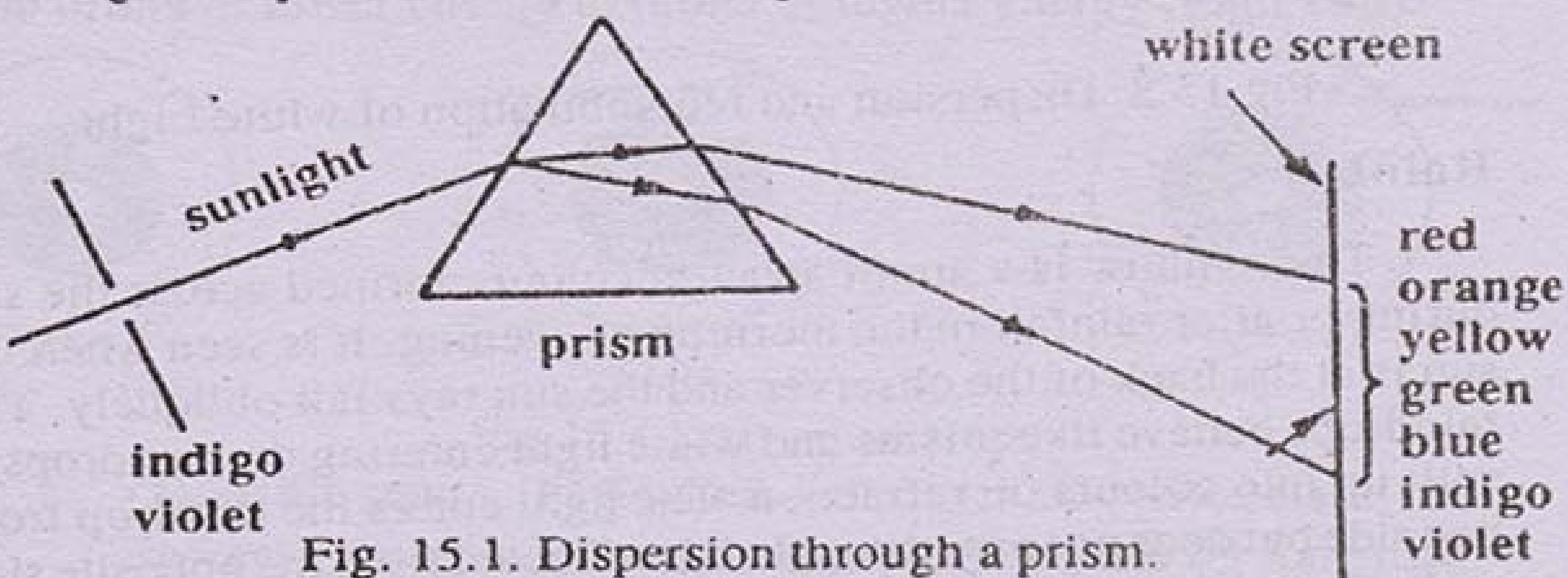


Fig. 15.1. Dispersion through a prism.

A beam of sunlight was allowed to enter a darkroom through a hole in the window. A prism was placed in the path of the beam of light A band of various colours was produced on the wall (Fig. 15.1). This band of colour is known as a *solar spectrum* and the phenomenon of spreading out of light into its constituent colours is called *dispersion of light*. How does it happen? The answer is simple. We know that the amount of refraction of the waves depends on their frequencies. This experiment shows that sunlight is a combination of various waves. When they pass through a glass prism refraction takes place. The waves of higher frequencies bend more than those of lower frequencies. Thus different waves with different wavelengths and frequencies in the visible light region would separate in the shape of a spectrum of colours. The shortest wavelength visible to the human eye is violet with 25,000 waves per cm (i.e. 0.4 μm) and its deviation is the greatest. The longest visible wavelength is red colour with about 14,286 waves per cm (i.e. 0.7 μm) and its deviation is the least.

When Newton place a second prism in the spectrum a patch of white light was obtained in place of a spectrum. The second prism refracts the colours in the opposite direction and so they recombine to form white light (Fig. 15.2). Thus in this way Newton proved that white light was made up of different colours.

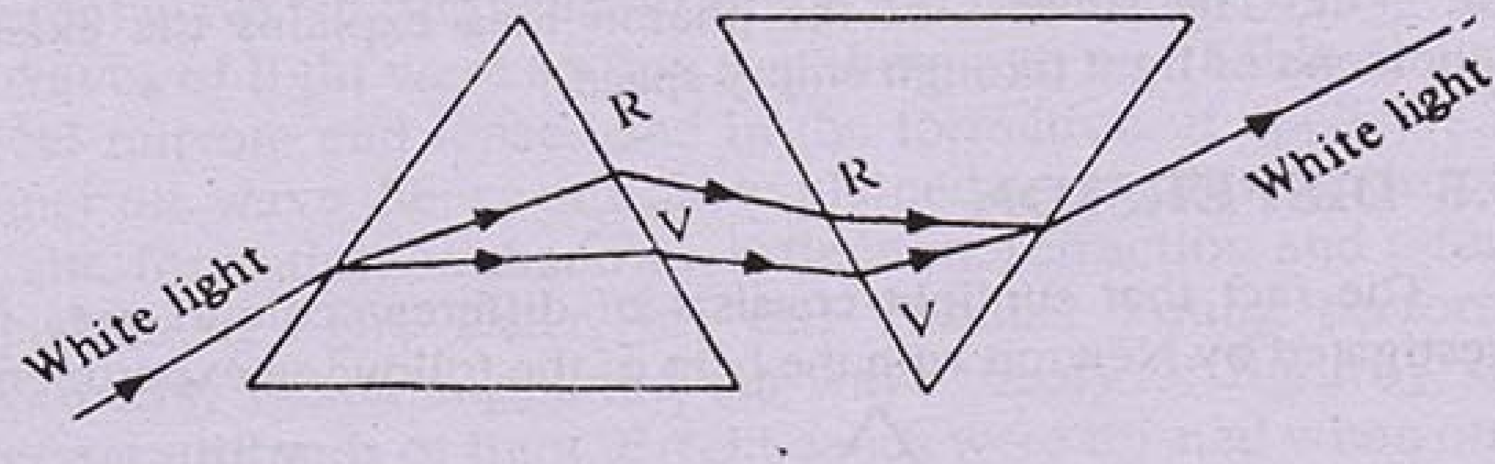


Fig 15.2 Dispersion and recombination of white Light

Rainbows

The rainbow is a arc of spectral colours formed across the sky during or after rainfall in the morning or evening. It is seen when the sun is at the back of the observer and the sun rays fall obliquely. The raindrops behave like prisms and white light entering the raindrops is split up into colours on refraction. The light enters the raindrop from one side but does not pass through. It is reflected from the opposite side and re-emerges through the side it entered but at an angle of about 42° to its original direction. Fig. 15.3, shows the paths of the extreme (violet and red)rays. The eye if placed at the position shown in figure, sees an arc with the red on top and the violet below. Some times the light is further reflected before emerging and this gives rise to a secondary rainbow seen above the first one.

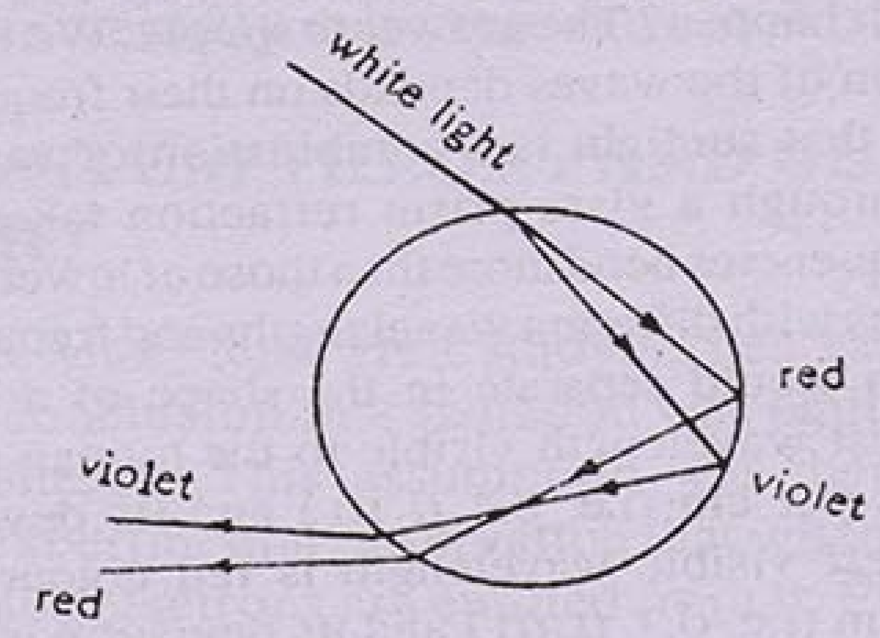
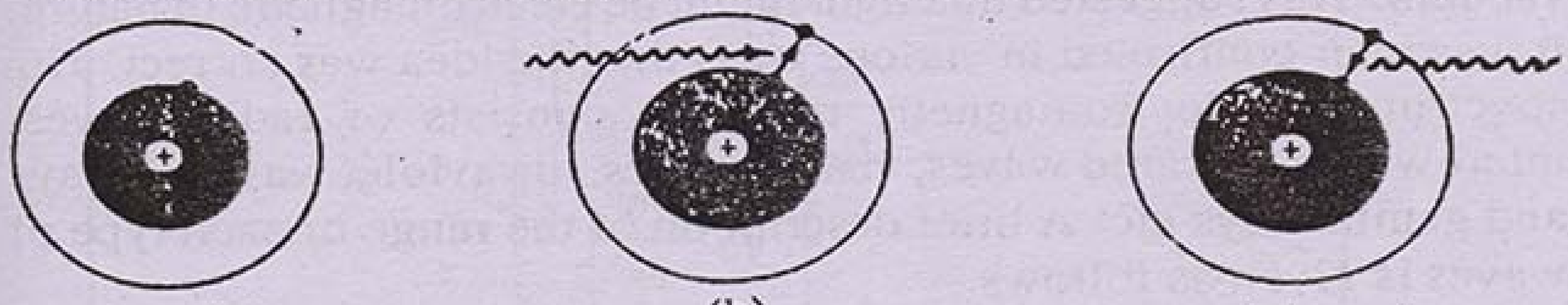


Fig. 15.3 Formation of Rainbow.

15.6 EMISSION OF LIGHT BY ATOMS

Besides the sunlight we obtain light from many sources. When solid is heated, its temperature rises and it starts to emit heat. At sufficiently high temperature, it becomes red hot. At still higher tempeature it becomes white hot. When gases are burnt, they also produce light. How is light produced? Since light consists of electromagnetic waves and all substances are made up of atoms, we should seek how the emission of light takes place in atoms. According to Bohr's atomic model electrons move round the nucleus only in certain discrete orbits without radiating energy. The energy of the electron in each orbit is well defined (Fig.15.4a). The orbit closest to the nucleus corresponds to the lowest energy. To explain the emission of light by atoms, Bohr proposed that when atoms are excited, electrons jump from orbits of lower energy to orbits of higher energy (Fig.15.4b).



(a) (b) (c)

Fig 15.4 State of an atom

(a) Stable state (b) Excited state (c) Deexcited state

However, they cannot remain in the excited state for long time. When the electron jumps back to orbits corresponding to lower energy, it emits energy in the form of photon (Fig.15.4c). The energy possessed by this photon precisely equals the difference of energy between the two orbits. The frequency of photon is proportional to the energy difference.

The return of excited atoms to their stable state results in the emission of light whose frequency depends on the atom. The frequencies of radiation emitted together form a spectrum called the bright line spectrum. Each atom of an element has its own characteristic spectrum. Light emitted from sodium bulb, mercury bulb and flourescent tube is due to this mechanism.

15.7 THE SPECTRUM

As discussed above when the dispersion of sunlight or white light is produced with the help of a prism, a band of different colours is

observed consisting of red, orange, yellow, green, blue, indigo and violet colour. A band of colour formed in this way is called a spectrum. Each colour in this spectrum represents a certain range of frequency and wavelength. The visible spectrum range from about 400 nm in the deep violet to about 750 nm in the deep red (1 n.m = 10^{-9}m).

The Electromagnetic Spectrum

A Scottish physicist, James Clerk Maxwell (1831-79), working on electromagnetism predicted that electrical oscillations would generate electromagnetic waves. He also derived a formula for the speed of electromagnetic waves in terms of electric and magnetic quantities. When these quantities were measured and the speed of electromagnetic waves was calculated it was discovered that the velocity of electromagnetic waves is same as the velocity of light in vacuum. This suggested that light might be electromagnetic in nature. It was later confirmed in various ways that the idea was correct. The spectrum of electromagnetic radiation consists of radio waves, microwaves, infrared waves, visible waves, ultraviolet waves, X-rays and gamma rays etc. A brief description of the range of each type of waves is given as follows.

Radio Waves are electromagnetic waves with a large range of wavelengths from a few millimetres to several metres.

Microwaves are radio waves of shorter wavelengths between 1 mm and 300 mm. Microwaves are used in radar and microwave ovens.

Infrared waves are also called heat waves, these waves are radiated by hot bodies at different temperatures. The earth's atmosphere is at a mean temperature of 250 K and radiates infrared waves with a wavelength having a mean value of 10 micrometres or 10^{-5}m($1\,\mu$m$=10^{-6}$m).

Visible Waves have a wavelength range between 400 and 700 nanometres. The peak of the solar radiation is at a wavelength of about 550 nm. The human eye is most sensitive to this wavelength.

Ultra Violet waves, their wavelength ranges from 380 nm down to 60 nm. These are emitted by hotter stars having a mean temperature greater than 25000°C.

X-rays wavelength ranges from 1.0 nm to 0.01 nm.

Gamma rays their, wavelength is less than 10^{-11} m. They are emitted by the nucleus of certain radio-active substances. They are also released during certain nuclear reactions . These wavelength ranges are not exact and the values given here are only for a general guide line. (Fig. 15.5).

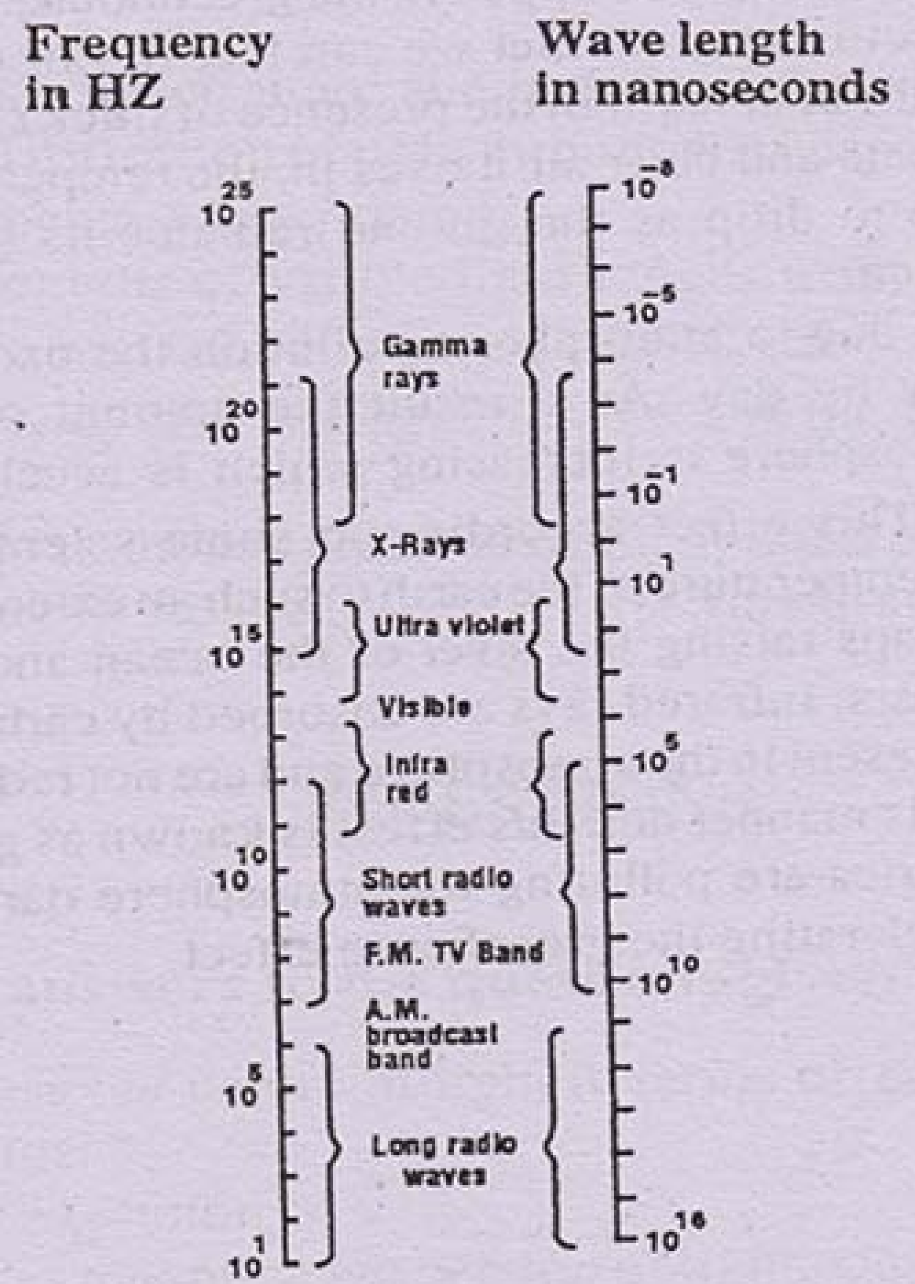


Fig. 15.5 Electromagnetic spectrum with various ranges

15.8 GREENHOUSE EFFECT AND GREEN HOUSES

The major portion of energy which the earth receives from the sun is absorbed by the atmosphere and surface of the earth giving rise to heating of, gases of the atmosphere and the water of the surface. The earth then becomes warm and radiates energy mostly in the form of

infrared waves of longer wavelength. Dust and gas molecules scatter some light back to space. The small amount of water vapour and carbon dioxide in the atmospbere is transparent to visible light but opaque to infrared radiation. This results in the trapping of heat energy in the lower part of the atmosphere. An analogous effect takes place in greenhouses where the glass of a greenhouse allows light to pass through and heat the enclosed ground. Since glass is relatively opaque to infrared radiation, the heat is trapped in the greenhouse. Sometimes the greenhouse effect is so great that we can feel it. For example, in winter it is much warmer at night in the presence of thick cloud layers. If such layers dissipate and clear air moves in, the temperature of the surrounding is sure to drop as the ground radiates its heat energy unhindered into space.

We know that due to atmospheric pollution the ozone layer is being damaged day by day. As a result the amount of radiation penetrating the atmosphere is increasing which is accelerating the greenhouse effect. This effect according to some scientists, might ultimately raise the temperature of the earth to such an extent that it will melt the polar ice-caps raising the level of the ocean and of course drowning all the coasts. Infrared rays are absorbed by carbon dioxide and water vapours present in the atmosphere and are not radiated back. Heat is trapped in this manner and this effect is known as *greenhouse effect*. Heavy industries are polluting the atmosphere damaging the ozone layer and accelerating the greenhouse effect.

SUMMARY

— Light is a form of energy that radiates from atoms when their electrons jump from higher orbit to lower one.
— Light is emitted or absorbed in small bundles of energy called photons, which travel in straight lines by a wave like motion.
— Light sometime behaves as particles and sometimes as waves.
— The phenomenon of splitting white light into its colours is called dispersion.
— Rainbow is a spectral arc and is formed during rainfall due to the dispersion and reflection of light through the suspended water drops in the air which behave like prisms.
— The wavelengths of visible light ranges from 0.4 μm to 0.7 μm ($1 \mu m = 10^{-6}$ m) or 1/30,000 cm to 1/60,000 cm.
— The velocity of electromagnetic waves is the same as the velocity of light.
— The term greenhouse effect is often applied to the heating of the atmosphere due to the presence of water vapours and carbon dioxide.

QUESTIONS

15.1 Write answers to the questions given below:

(i) Discuss the nature of light in terms of corpuscular and wave theories.
(ii) What is photon.
(iii) What is meant by Dispersion of light. Describe an experiment to show that light consists of seven colours.
(iv) How is rainbow formed.

15.2 Fill in the blanks

(i) Light consists of minute particles called ________ according to the Corpuscular theory of light.
(ii) Light is a form of ________ which propagates as trains of ________ according to wave theory of light.
(iii) According to quantum theory light is supposed to be a form of ________ which is absorbed and emitted in________.

(iv) The speed of light in air is ________ than the speed of light in water.

(v) ________ light is composed of lights of different ________.

(vi) Splitting up of white light into several colours is called ________.

(vii) According to ________ light is propagated in space in the form of waves.

(viii) Electromagnetic waves differ in frequency and in ________.

(ix) Light is emitted by ________ atoms.

(x) An electron may absorb energy and be lifted to a higher ________.

15.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) The speed of light is ________ m/s.

(a) 3×10^6 (b) 1.86×10^6 (c) 3×10^8 (d) 3×10^{10}

(ii) Small drops of rain water disperse sunlight into different colours. This is called ________.

(a) Dispersion (b) Interference

(iii) According to Quantum theory of light photons are ________.

(a) Waves (b) Electromagnetic waves

(c) Energy packets (d) particles.

(iv) According to Hygen's wave theory, light propagates in the shape of ________.

(a) Photons (b) Waves (c) Particles

(v) We use a ________ to disperse white light into different colours.

(a) Convex mirror (b) prism

(c) convex lens (d) concave mirror.

(vi) The radiation which produces the sensation of heat is called ________.

(a) Invisible light (b) Ultra-Violet rays

(c) Infrared rays (d) Visible light.

(vii) Electromagnetic waves carries ________.

(a) Wavelength (b) Frequency (c) Charge (d) Energy

(viii) Which of the following are not electromagnetic in nature.

(a) Infrared rays (b) Ultra-violet rays

(c) Radar waves (d) Sound waves.

(ix) A ray of white light is shown onto a glass prism. The light can not be __________.

(a) deviated (b) dispersed (c) focused (d) refracted.

(x) Which of the following examples of electromagnetic radiation has the shortest wavelength?

(a) Radio waves (b) Infrared rays

(c) Ultra-violet rays (d) X-rays

15.4 Pick out true and false from the following sentences.

(i) Infrared radiation stimulates the formation of vitamin D.

(ii) Due to atmospheric pollution the oxygen layer is being damaged day by day.

(iii) The frequency of visible light waves ranges from to $0.1\mu m$ $0.3\mu m$.

(iv) Rainbow appears in the sky on the side opposite to where the sun is located.

(v) White light is refracted while passing through a prism. The amount of refraction is not the same for all the colours composing the white light.

(vi) The frequency of a photon of violet light is double than the frequency of a photon of red light.

CHAPTER - 16

ELECTRICITY

16.1 INTRODUCTION

you have already learnt in the lower classes that charge is produced when an ebonite rod is rubbed with a woolen cloth or a glass rod with a silk cloth. You have also studied the properties of the charges. In this chapter we shall recapitulate the previous knowledge and shall further study different phenomena such as, the force between the charges, the field due to the charges, the potential and the intensity of the field due to the charges i.e., the electric field etc.

In the case of moving charges we will learn about the terms like,. current, resistance, potential difference and a law that connects them. Also we shall study the current measuring instruments and the devices that store charges and the difference between direct and alternating current. At the end a brief description about the house-hold wiring and the safety precautions will be given.

16.2 ELECTRICAL NATURE OF MATTER

You have already studied that an atom, the fundamental constituent of matter, is composed of three elementary particles called neutron, proton and electron.

Proton is positively charged particle and is found in the nucleus of an atom. It is considered to be a heavy particle because it is about 18 times more massive than electron.

Neutron is a neutral particle i.e. it has no charge on it. Neutron is confined in the nucleus. It's mass is approximately equal to that of a proton.

Electron is a negatively charged particle. The amount of charge on it is equal to that on a proton. An atom may contain one or more than one electrons. Since an atom in normal state is a neutral particle, the number of protons in a nucleus is equal to the number of electrons revolving around it.

It may be pointed out that although the electron and proton are different in mass but they have exactly the same amount of electric charge.

All material objects are composed of atoms. When one or more than one electrons are removed from an atom it becomes positively charged particle because of excess protons in the nucleus; on the other hand if electrons are added in an atom it becomes negatively charged due to excess of electrons. In the first case the number of protons in the nucleus are more than the number of electrons revolving around the nucleus, while in the second case number of electrons around the nucleus are more than the number of protons in the nucleus. This is the basic reason why positive and negative charges appear over an object.

16.3 INSULATORS AND CONDUCTORS

Suppose that we have two metallic sphere one highly charged and the other without any charge and we connect the two spheres by placing a wooden stick, a plastic stick, a glass rod and a rubber strip one after the other between them. If we test whether any charge has moved from the charged sphere to the uncharged one through any of the above materials it is found that no appreciable amount of charge has passed through any of the material to the uncharged sphere. Now if we repeat this experiment by placing an iron wire between the two spheres to connect them and test whether any charge has passed through the iron wire to reach the uncharged sphere it will be found that an appreciated amount of charge has reached the uncharged sphere after passing through the iron wire. The same result is observed if we use copper wire, aluminium wire or any other metallic wire to connect the two spheres. Thus we see that there are two categories of material. Those material objects which do not allow the charge to pass through them belong to one category and are called insulators or non-conductors. Wood, plastics, rubber, etc belong to this category. On the other hand, those material objects which allow the charge to pass through them are called conductors. Materials like copper, iron, aluminium, gold, silver etc, belong to this category.

The electrons in an atom of an insulator are bound tightly with the nuclei and thus charge cannot pass through them. In a conductor some of the electrons are loosely bound and can move about freely within the material. When a positively charged object is brought close to or touches a conductor, the free electrons move quickly towards this positive charge. On the other hand the free electrons move swiftly away from a negative charge if is brought close to them.

There is a third class of materials whose properties lie between those of insulators and conductors. These materials are called semi-conductors and will be dealt with later.

Coulomb's law

When one charged body is brought near another charged body, either a repulsion or an attraction is observed between them. This shows that some kind of force is operated between them. This force is known as an electrostatic force. The force is found to be an attractive. When a positively charged body is brought near a negatively charged body (Fig. 16.1.a), and it is found to be repulsive if the two bodies have the same charge on them (Fig. 16.1.b).

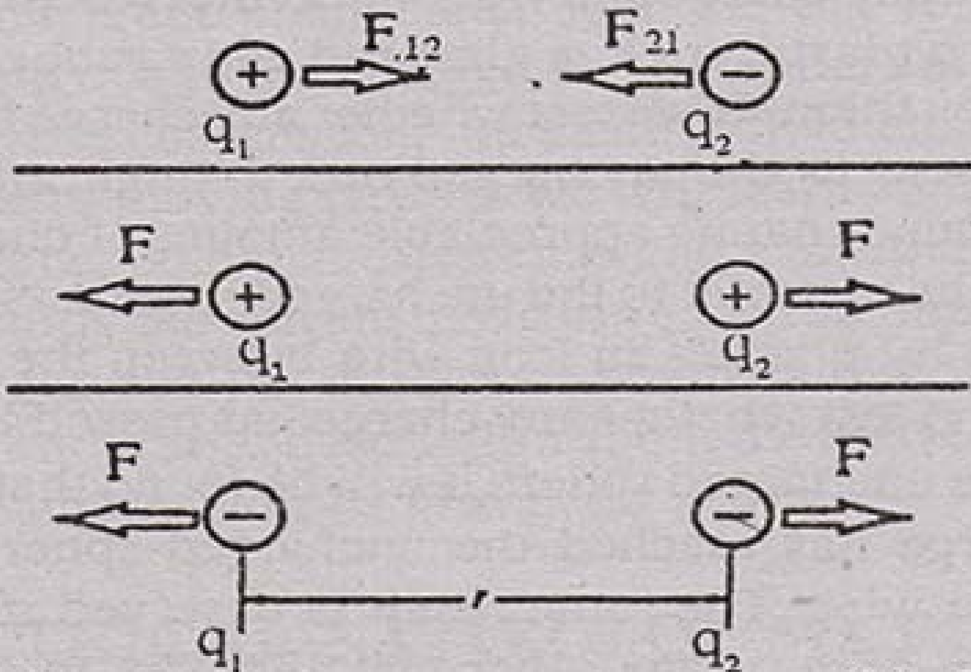


Fig. 16.1(a) Unlike charges attract each other. (b) Like charges repel each other.

If the physical sizes of these charge are very small, compared to the distance of separation between them, then the charged objects can be regarded as points. The point charges have magnitudes q_1 and q_2. If the charges have unlike signs in Fig. 16.1.a, each charge is attracted to the other by a force that is directed along the line joining them. In Fig. 16.1 (a) $+ F_{12}$ is the electrostatic force exerted on charge 1 by charge 2 and $- F_{21}$ is the electrostatic force exerted on charge 2 by charge 1. If the charges have the same sign (either both positive or both negative), as in Fig. 16.1 (b.c), then each charge is repelled from the other. The repulsive force, like the attractive forces, acts along the line joining the charge. The two forces, whether attractive or repulsive, are equal in magnitude but opposite in direction. These forces always exist as pairs, each one acting on a different object.

The French Scientist Charles Augustin De Coulomb experimentally investigated these observations and came up with a law known as Coulombs law. It states that the magnitude of the electrostatic force between two point charges is directly proportional to the product of the magnitude of the charges and inversely proportional to the square of the distance between them. If q_1 and q_2 are the magnitudes of two charges separated by a distance r, the electrostatic force. (F) between them according to this law is

$$F \propto \frac{q_1 q_2}{r^2}$$

or

$$F = k\frac{q_1 q_2}{r^2} \quad \text{-------------} \quad (16.1)$$

where k is a constant of proportionality whose value in SI units is 8.99×10^9 or $9 \times 10^9\ Nm^2\ C^{-2}$. $\left(\frac{\text{newton - metre}^2}{\text{coulomb}^2}\right)$

The constant k is commonly expressed in terms of an other constant ϵ_o (epsilon), by writing

$$k = \frac{1}{4\pi\epsilon_o}$$

Where ϵ_o is called the permittivity of free space.

Unit of Charge

Since the units of distance and force are known, the Eq. 16.1 can be used to define the unit of electric charge.

The unit of charge in SI system is the coulomb. One coulomb (1C) of charge being that quantity of charge which when placed one metre from an identical charge in vacuum repels it with a force of 8.99×10^9 newton (N).

$$F = k\frac{q_1 q_2}{r^2}$$

$$8.99 \times 10^9 = k\frac{(1C)^2}{(1m)^2}$$

$$k = 8.99 \times 10^9\ \frac{N - m^2}{C^2} \cong 9 \times 10^9\ N - m^2\ c^{-2}$$

If the medium in which these charges are placed is not vaccum but a material medium then we have to include the properties of the medium represented by $\in r$, called the relative permittivity of the medium.

Now equation 16.1 can be written as

$$F = \frac{1}{4\pi \in_{\circ} \in_{r}} \times \frac{q_1 q_2}{r^2}$$

Following are the submultiples of coulomb

1 milli coulomb (1mc) = 10^{-3} coulomb

1 micro coulomb ($1\mu C$) = 10^{-6} coulomb

1 micro micro coulomb ($1\mu\mu C$) = 10^{-12} coulomb

16.4 ELECTRIC FIELD

Coulomb's law enables us to calculate the magnitude as well as the direction of the electriostatic force. This law is limited to describing only interactions between two point charges. But how is the force transmitted from one charge to another without their coming in contact. This idea of a force "action at a distance" is a difficult concept to comprehend. This conceptional difficulty can be overcome with the idea of a field which was developed by Michael Faraday. According to the field concept the interaction between two charges q and $q_{\circ}$ separated by a distance r is explained in the following way.

The charge q produces an electric field that permeates the entire space surrouding the charge. We define an electric field due to a charge as the space around that charge in which its influence would be felt by another charge (Fig. 16.2.a). It is useful to regard $q_{\circ}$ as a test charge for determining the extent to which a charge q generates a force on $q_{\circ}$. Idealy the test charge is so small that it does not disturb other field of the charge in which it is present otherwise it would not detect the original field. The test charge is standardised as a unit positive charge. The interaction between q and $q_{\circ}$ is a two step process. (i) The charge q produces a field (E) and (ii) The field E interacts with $q_{\circ}$ to produce a force F on $q_{\circ}$. These two steps are shown in (Fig. 16.2.b).

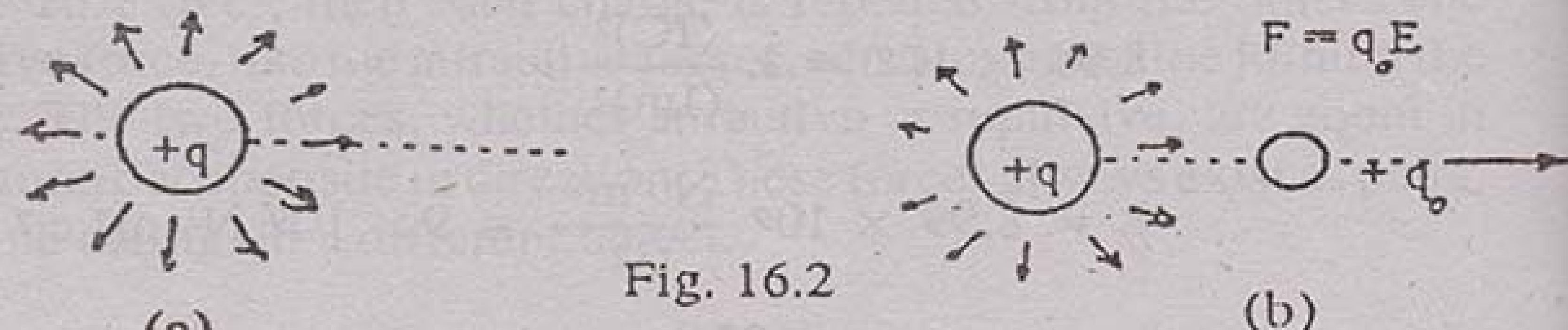


Fig. 16.2

The two steps are reversible i.e, charge q_o may produce a field and if a charge is brought into the field of q_o at a distance r, it will feel a force which is equal in magnitude and opposite in direction.

From Coulomb's law we know that the magnitude of the force F depends upon the separation r between the two charge. In terms of the field concept we say that the strength of the field is very high near charge q and decreases as $\frac{1}{r^2}$. From here we can define the electric field strength or electric field intensity E due to a charge q at a distance r from it as:

$$E = \frac{F}{q_o} = \frac{q}{4\pi \epsilon_o r^2} \quad \text{----------} \quad (16.2)$$

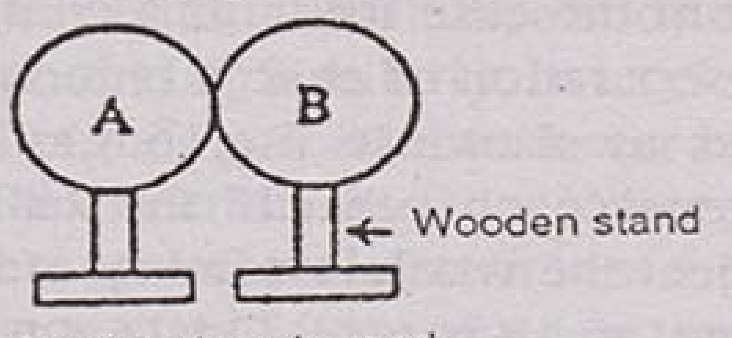
where F is the electrostactic force due to q on q_o. Thus the electric intensity E at any point surrounding the charge q is defined as the force per unit positive charge. As E is a vector quantity its direction is the same as that of the force that will act on a unit positve charge or test charge in the field. Its SI units are newtons per coulomb (NC^{-1}).

16.5 ELECTROSTATIC INDUCTION

A piece of paper when brought near a charged body, it is attracted by it. The attraction of neutral body by a charged body can be explained on the basis of structure of atom.

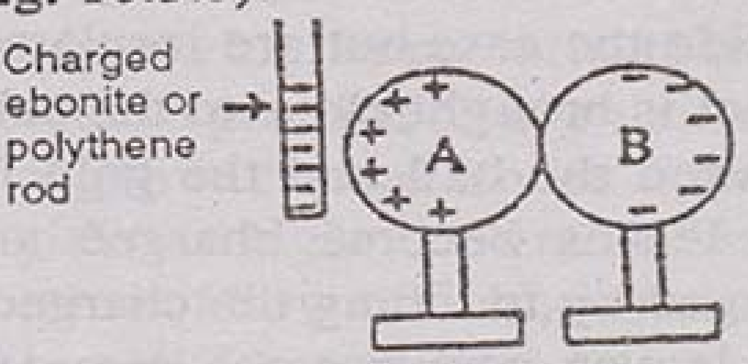
(i) Two metal spheres A and B fitted with wooden stands are placed together so that they touch one another and thus form in effect a single conductor as shown in (Fig. 16.3.a).

(ii) A negatively charged ebonite rod is now brought near the sphere A such that it does not touch the sphere A. As a result positive charge induced on sphere A and negative charge on the sphere B as shown in (Fig. 16.3.b).



Insulated uncharged conductors in contact

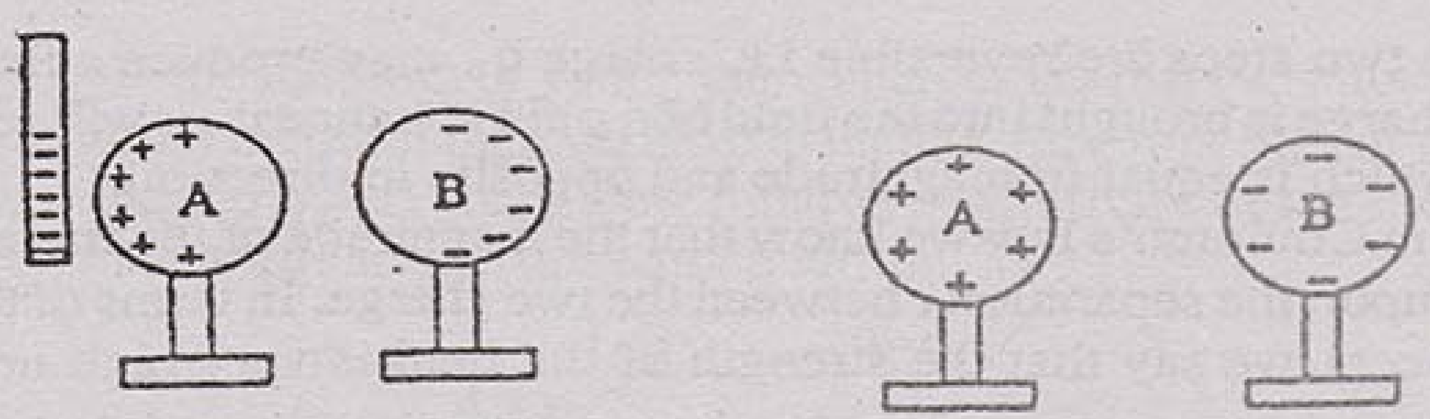
(a)



+ve and −ve charges are induced on A and B respectively

(b)

Fig. 16.3



A and B separated in presence of inducing charge

(c)

Inducing charge removed A and B have opposite charges

(d)

Fig. 16.3

(iii) Still keeping the charged rod in the same position, the sphere B is moved-away from sphere A at a short distance as shown in (Fig. 16.3.c).

(iv) The negatively charged rod is now removed as shown in (Fig. 16.3.d) and then the sphere A and B are tested for charge.

In order to test the nature of charges on sphere A and B an electroscope is used. It is observed that A is positively charged and B is negatively charged.

If whole experiment is carried out again by using positively charged rod as the inducing charge, the charge on A and B will beinversed i.e. the sphere A will be negatively charged and the sphere B will be positively charged. Separation of charges by this phenomenon is known as electrostatic induction.

Gold Leaf Electroscope

A gold leaf electroscope is a device that can be used for detecting charge. It consist of a glass case that contains two thin leaves of gold which are capable to diverge shown in Fig.16.4. The leaves are connected through a conductor to a metal ball or disk outside the case,but are insulated from the case itself .If a charged object is brought close to the ball, a separation of charge is induced between the ball and the gold leaves as shown in Fig.16.4.a. The two leaves become charged and repel each other.If the ball is charged by touching the charged object the whole assembly of ball and leaves acquires the same charge as shown in Fig.16.4.b. In either case greater the amount of charge greater would be the divergences in leaves.

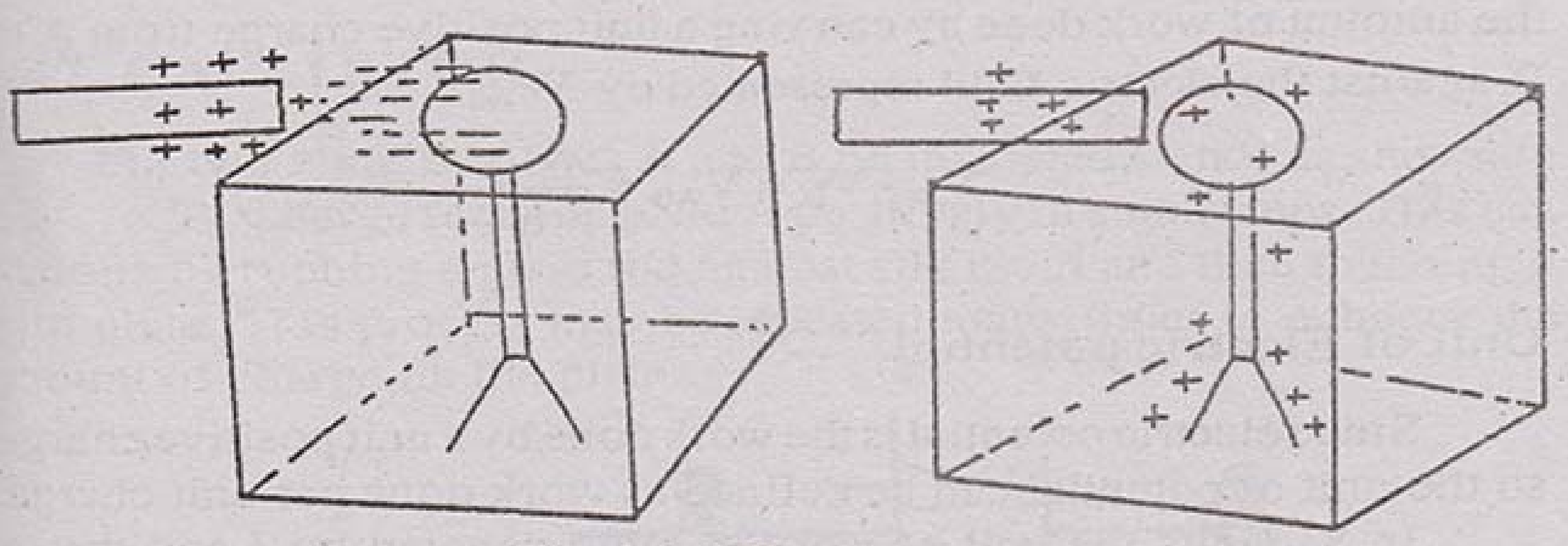


Fig. 16.4 Electroscope

Whether a body is positively or negatively charged it can be determined by bringing it near a charged electroscope.

16.6 ELECTROSTATIC POTENTIAL

In raising a body above the surface of the earth, the work has to be done against the earth's gravitation and it increases its gravitational potential energy. Now if the body is released from this height, it falls towards the earth. This implies that a body moves from the higher gravitational potential towards lower one.

If a positive charge is placed in an electric field, then it will move along the direction of the electric field. On the other hand, if the positive charge is made to move opposite to the direction of the electric field, then a force has to be applied on this charge and work has to be done on it during its motion. The charge thus acquires an electro-static potential energy.

This accumulation or potential energy is termed as the increase of electreic potential of the charge. If this charge is set free then it will move from a place of higher potential to lower potential and energy equal to the difference of potential will be released.

In gravitational field, a body moves from a higher potential to the lower potential. Similarly, in an electric field a positive charge moves from a higher potential to lower but a negative charge moves in the opposite direction i.e. from lower potential to a higher potential. Just as difference in level determines the direction of flow of fluid, pressure difference that of flow of a gas, temperature difference that of flow of heat, similarly, the electric potential difference determines the flow of electric charge. Consider two points A and B at different potentials in an electric field. The potential difference between A and B is equal to

the amount of work done by carrying a unit positive charge from A to B against the electric field represented by W_{AB}.

$$V_A - V_B = V_{AB} = \frac{W_{AB}}{q}$$

Unit of Electric potential

Since electric potential is the work done by a unit positive charge, so the unit of potential can be defined as work done per unit charge.

In SI Units the unit of work is Joule denoted by J and that of charge is Coulomb denoted by C so the unit of potential in SI. units is Joule per Coulomb or JC^{-1}. It is also called volt denoted by V. Hence,

$$\text{Volt} = \frac{\text{Joule}}{\text{coulomb}}$$

In an electric field, the potential between two points is said to be 1V if the amount of work done by 1 coulomb charge (6.25×10^{18} electrons) in moving from one point to another is one joule.

Example 16.1

The potential difference between two points is 110V. If an unknown charge is moved between these points, the amount of work done is 550J. Find the amount of charge.

Solution.

Given that the potential difference between two points = 110V
The amount of work done to move the charge = 55oJ

Since

$$\text{Potential difference} = \frac{\text{work done}}{\text{charge}}$$

$$110V = \frac{550J}{q}$$

or

$$q = \frac{550J}{110V} = 5C$$

Hence the unknown charge is 5C

16.7 CAPACITOR

Fix two metallic plates A and B on insulated stands as shown in (Fig. 16.5). Charge metallic plate A positively in small steps. This can be done by rubbing a glass rod against silk cloth and then touching it with plate. This process may be repeated many times to enhance the amount of charge on the plate.

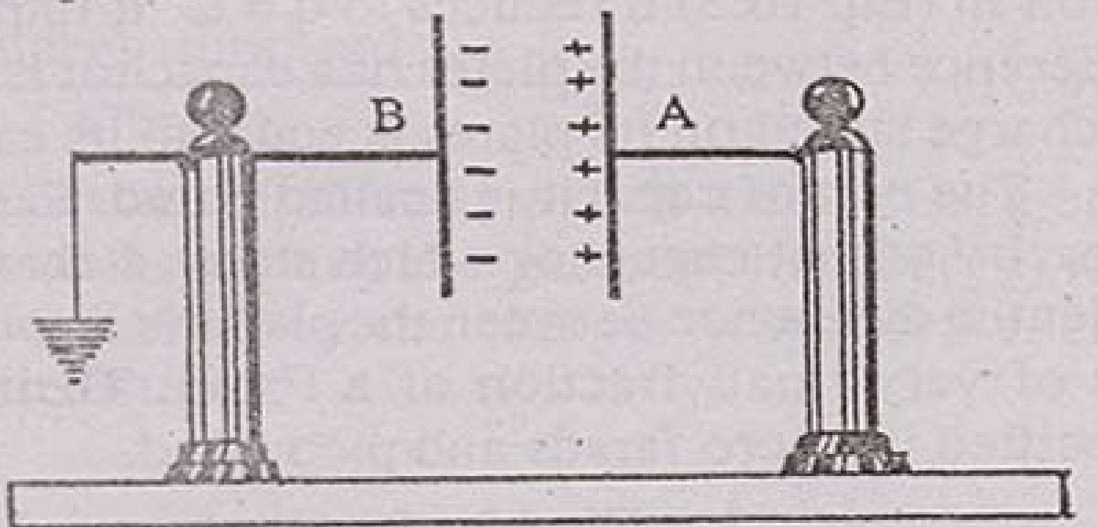


Fig. 16.5

This increases its potential. It must be kept in mind that this process can not be continued indefinitely. After a while the plate A can not be charged any more. The amount of work done in carrying positive charge to the plate is converted into electrical potential energy. The other plate will have a negative charge induced on it and these positive and negative charges hold each other. The system of these two plates is called a Capacitor. It is used to store charge and electrical energy. The capacitor showin in (Fig. 16.5) is called parallel plate capacitor as the charge is stored on parallel plate. A capacitor can be of other shape as well. If an insulator such as glass, mica etc. is introduced between the plates, its capacity to store charge increases. Experiments have also established the facts that increase in the area of the plates of a capacitor or the reduction in the distance between them also enhances its capacity to store charge.

The stored electrical energy in a capacitor can be utilized according to the need. It is also a very important component of electrical appliances.

Unit of Capacity of a Capacitor

When charge q is transferred on one of the plates of a capacitor, the potential difference V between the plates also increases. In other words, the charge q on the plate of capacitor is directly proportional to the electric potential difference V between them i.e.

$$q \propto V$$

or

$$q = CV \qquad \text{------------} \qquad (16.3)$$

where C is a constant whose value depends upon the area of the plates, the distance between the plates and the medium between them. The constant C is called the capacity of the capacitor. Substituting V with 1 volt in (Eq. 16.3) it reduces to $q = C$. It implies that if the potential difference between the plates of a capacitor is 1 volt then the quantity of charge stored on its plates is equal to its capacity.

The unit of capacity is called Farad. Farad (F) is defined as the capacity of that capacitor which stores a charge of 1 Coulomb if the potential difference between the plates is 1 volt. Generally, Capacitors are of very small fraction of a Farad. Their capacities are usually specified in micro farads and pico farad.

$1\ \mu$ F(micro farad) $= 10^{-6}$ farad

$1\ \mu\mu$ F(Pico farad) $= 10^{-12}$ farad

Example 16.2

The potential difference between the parallel plates of a capacitor is 750 V. If on each plate there is a charge of 0.006C, find the capacity of the capacitor.

Solution

The potential difference between the plates $= V = 750$ V

Amount of charge, q on each plate $= q = 0.006$ coulomb

Capacity C of the capacitor = ?

By using (Eq. 16.3)

$$0.006 \text{ coulomb} = C \times 750\text{V}$$

or

$$C = \frac{0.006 \text{ coulomb}}{750 \text{ V}} = 8\ \mu\text{F}.$$

16.8 ELECTROMOTIVE FORCE (e.m.f.)

When a source of electrical energy, such as a cell, is connected to an electric circuit it produces a current which passes through all the components in the circuit and itself. The energy required to drive the charge around the circuit is called electromotive force and is defined as the potential energy applied to each unit of charge. The unit of e.m.f. is volt

$$\text{e.m.f.} = \frac{\text{Energy supplied}}{\text{Charge}} \quad \text{------------} \quad (16.4)$$

The difference between P.D and e.m.f. is that a source of e.m.f. in a circuit does work on the moving charges (supplies energy) whereas P.D. refers to the work done by the charges (expended energy) in passing through the resistance.

Example 16.3

Calculate the potential difference between two points A and B if it requires 9×10^{-4} J of external work to move a charge of $+5\,\mu$C from A to B which is at higher potential?

Solution

Charge $\quad q_o = +5\,\mu C = +5 \times 10^{-6}$ C

$$W_{AB} = 9 \times 10^{-4} \text{ J}$$

As $$\Delta V = \frac{W_{AB}}{q_o} = \frac{9\times10^{-4}\text{J}}{5\times10^{-6}\text{C}}$$

$$\Delta V = 1.8\times10^{2} \text{ J/C} = 180\text{V}$$

16.9 ELECTRIC CELLS

There are many ways of obtaining electricity. The electric cell is one of them. There are different ways of making cells. You need three things to make a cell. Two of these are electrodes. the third thing required is a container containing an electrolyte solution.

Primary cells

Such type of cells whose working does not depend upon any external source are called primary cells, e.g voltaic cell, daniell cell,

leclanche cell and dry cell. In these cells the current is produced by the chemical action between its various components. When the chemicals are exhaused they can not furnish any more current. The cell has got to be prepared afresh.

Voltaic cell

Place a strip of zinc and a strip of copper in a glass vessel filled with dilute sulphuric acid. The strips are called electrodes and the sulphuric acid solution is known as an electrolyte. Connect the two strips by a copper wire. An electric current will be set up in the conductor as indicated by the lighting of the bulb connected in the circuit. It can be shown by means of a sensitive electroscope that the zinc plate has a negative charge while the copper plate has a positive charge. The copper strip shown in (Fig. 16.6), is thus called the positive (+) electrode and the zinc strip is called the negative (-) electrode. Let us see how this happens. When the zinc strip is placed in the acid solution it reacts with the acid and begins to dissolve. Zinc atoms leave the strip and enter the solution. Each zinc atom that escapes from the strip leaves behind two electrons. Thus the zinc atom when in solution becomes a positive zinc ion (Zn^{++}) where an ion is an atom or

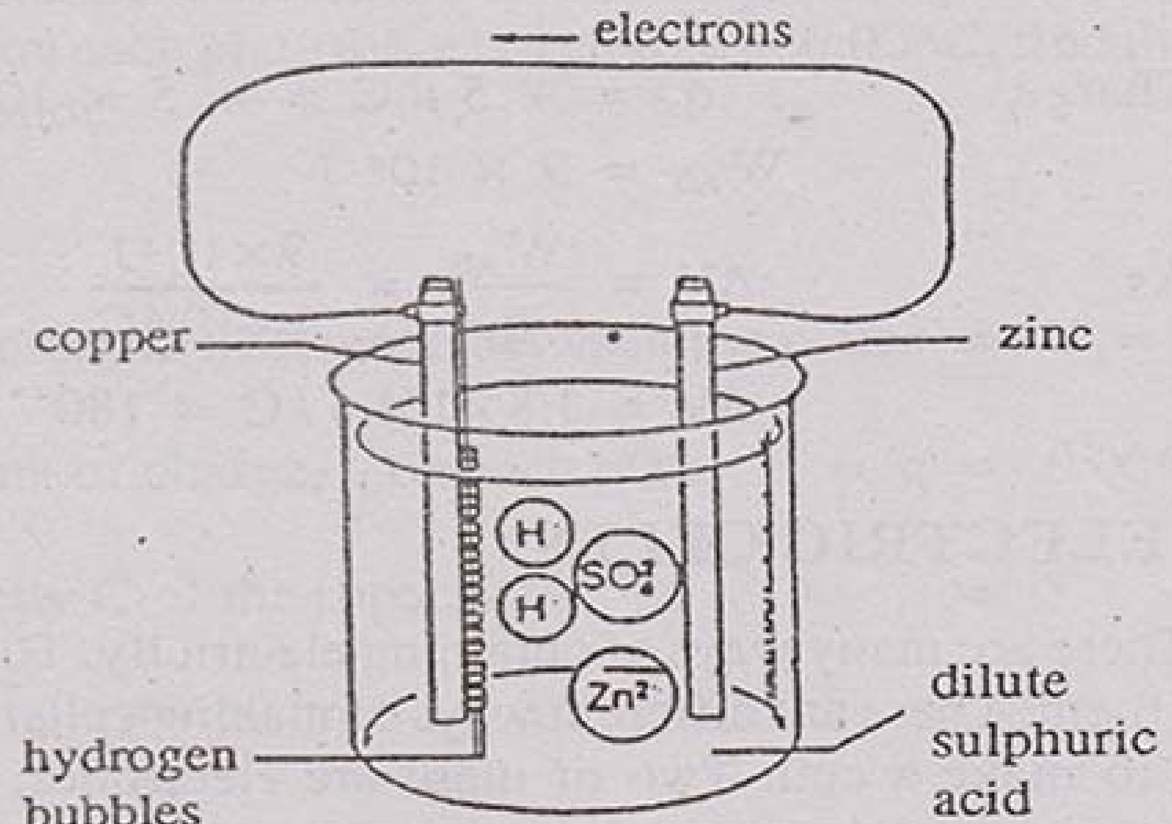


Fig. 16.6

a group of atoms that have a net positive or a net negative charge. At the same time, because of the electrons left behind, the zinc strip becomes negatively charged. This reaction can be written as

$$Zn \longrightarrow Zn^{++} + 2\,e^-$$

The sulphuric acid solution (H_2SO_4) consists of sulphate (SO_4^{-}) and hydrogen ions (H^+). The equation for this chemical reaction is given as

$$H_2SO_4 \longrightarrow 2H^+ + SO_4^-$$

As positive hydrogen ions reach the copper they remove electrons from the copper atom. The neutralized hydrogen atoms will form hydrogen molecules and then bubble off into the air. The copper strip having lost electrons becomes positively charged. When the positively charged copper strip is connected with the help of a wire to the negatively charged zinc strip, a current will start to pass from positive to negative electrode of the cell. This current goes on passing, so long as the zinc will continuously be dissolved and hydrogen continously liberated at the copper plate.

Voltaic cells though simple to construct are not much in common use because the current in them stops passing after some time. This is due to the fact that when a large amount of hydrogen drifts towards the copper plate, only a part of these hydrogen ion reach the copper plate as the hydrogen ions are not so rapidly neutralized. This causes an accumulation of hydrogen round the copper plate which prevents contact between the copper and the acid and so stops the flow of current. Such a defect of voltaic cells is known as polarisation. The polarisation may be prevented to some extent by giving the positive plate a rough surface from which the hydrogen bubbles breaks away more easily than from a smooth surface. The phenomenon of preventing from hydrogen layer being formed on the positive plate is known as depolarizing. The other defect which affects the performance of a voltaic cell is known as local action. It becomes prominent when the zinc plate has impurities such as carbon or iron. These impurities together with the zinc and acid form tiny cells at the surface of the zinc plate in a simple cell. Due to chemical reactions which take place within these tiny cells, zinc dissolves in the acid and bubbles of hydrogen are given off from the particles of impurities. This is known as local action and takes place even when the simple cell is not driving a current in the external cirucit. This can be removed by amalgamating the zinc plate. For this purpose the zinc plate is dipped into sulphuric acid and then rubbed with mercury which dissolves some of the zinc, forming a solution known as zinc amalgam which coats the plate. As the mercury does not dissolve the impurties, they are kept away from contact with the acid in the cell by the coating of amalgam over them.

The Daniell Cell

It is a modified form of the simple cell and is illustrated in (Fig. 16.7). It consists of a copper vessel containing saturated copper sulphate solution which acts as a depolariser and the vessel itself behaves as the positive plate. The negative plate is an amalgamated zinc rod which stands in a porous pot filled with dilute sulphuric acid. The porous pot in turn is placed at the centre of the copper vessel containing saturated copper sulphate solution. In this way the two fluids are in contact with each other through the pores of the porous pot.

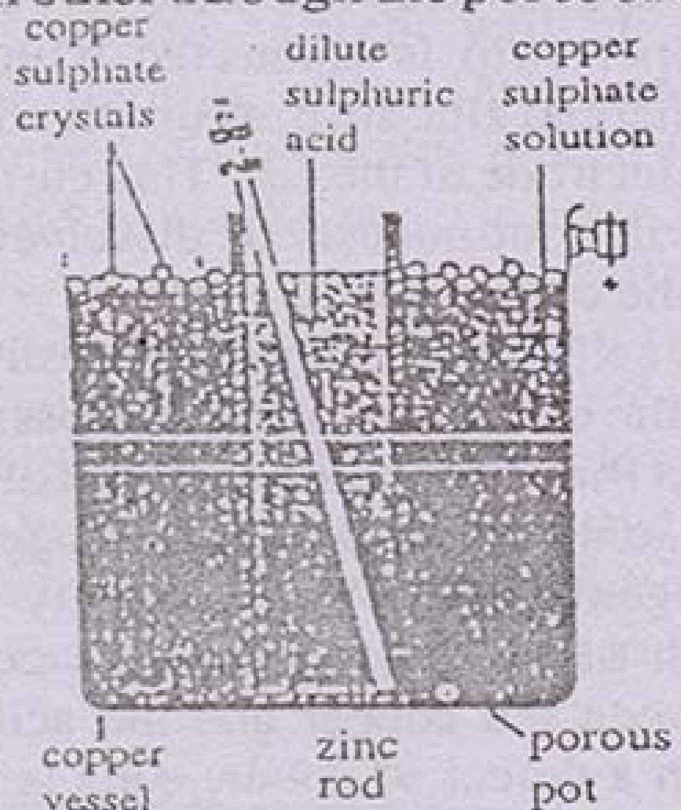


Fig. 16.7 Daniell Cell

When the cell is in use zinc gradually dissolves in the acid to form zinc ions leaving an excess of electrons on the zinc rod. An equivalent amount of hydrogen ions diffuses through the porous pot and enters the copper sulphate solution in which C_u^{++} and SO_4^- ions are present. The hydrogen ions (H^+) combine with the sulphate ions (SO_4^-) to form sulphuric acid (H_2SO_4) leaving behind copper ions which drift toward the copper vessel where they deposit their charge making it a positive terminal. In the above chemical reaction no hydrogen is liberated and hence the polarization is prevented.

When the cell is not in use, the porous should be removed and emptied otherwise the two liquids will gradually diffuse into one another. Also the zinc rod should occasionally be re-amalgamated otherwise there will be an excessive dissolving of zinc by the acid. The initial emf produced by the Daniell cell is about 1.1 volts.

The Leclanche Cell

The Leclanche cell is an improved version of the Daniell cell, and

has an emf of about 1.5 V which is relatively higher than that of the simple voltaic cell. The internal structure of this cell is illustrated in Fig. 16.8. The positive terminal of the cell is a carbon rod which is packed into a mixture of manganese dioxide and powdered carbon in a porous pot. The porous pot stands in a glass jar filled with a solution of ammonium chloride. A zinc rod dipped in this solution serves as the negative terminal of the cell. Hydrogen reaching the carbon rod is oxidized to water by the manganese dioxide.

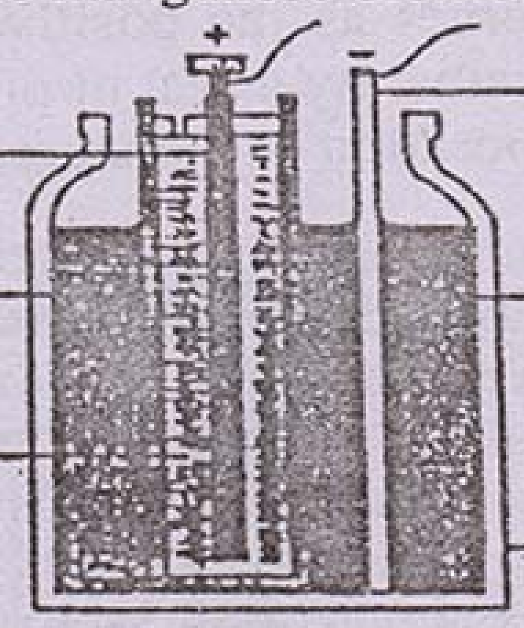


Fig. 16.8 Leclanche Cell

The chemical reaction which takes place between the ammonium chloride (NH_4CI), manganese dioxide (MnO_2) and zinc (Zn) results in the production of ammonia (NH_3), zinc chloride ($ZnCl_2$), water H_2O and magnanese oxide (MnO_3).

$$2MnO_2 + 2NH_4CI + Zn \longrightarrow Mn_2O_3 + H_2O + ZnCl_2 + 2NH_3$$

The essential electron exchange occurs when zinc metal (Zn) gives up electrons at the negative terminal to become zinc ions (Zn^{++})

$$Zn \rightarrow Zn^{++} + 2e$$

Charge electron flow through the external circuit to the carbon rod or positive terminal where they are taken up by the manganese. However, the depolarizing action takes place slowler than the rate at which hydrogen is liberated. Thus if the cell is used continuously polarization takes place and the emf of the cell drops quickly. Hence this type of cell can be used only in those cases when intermittent currents of short duration are required e.g. electric bells.

The Dry Cell

The dry cell Fig. 16.9 is a modified form of the Leclanche cell. The container is made of zinc and serves as the negative terminal. Inside it is a paste of ammonium chloride, manganese dioxide, powdered carbon, a little water and neutral filler such as flour. It is not perfectly dry but is called a dry cell because the ammonium chloride solution of the Leclanche cell is replaced by ammonium chloride paste. A central carbon rod serves as the positive terminal. The various components are held together by a seal of pitch or plastic which also prevents leakage or evaporation of moisture.

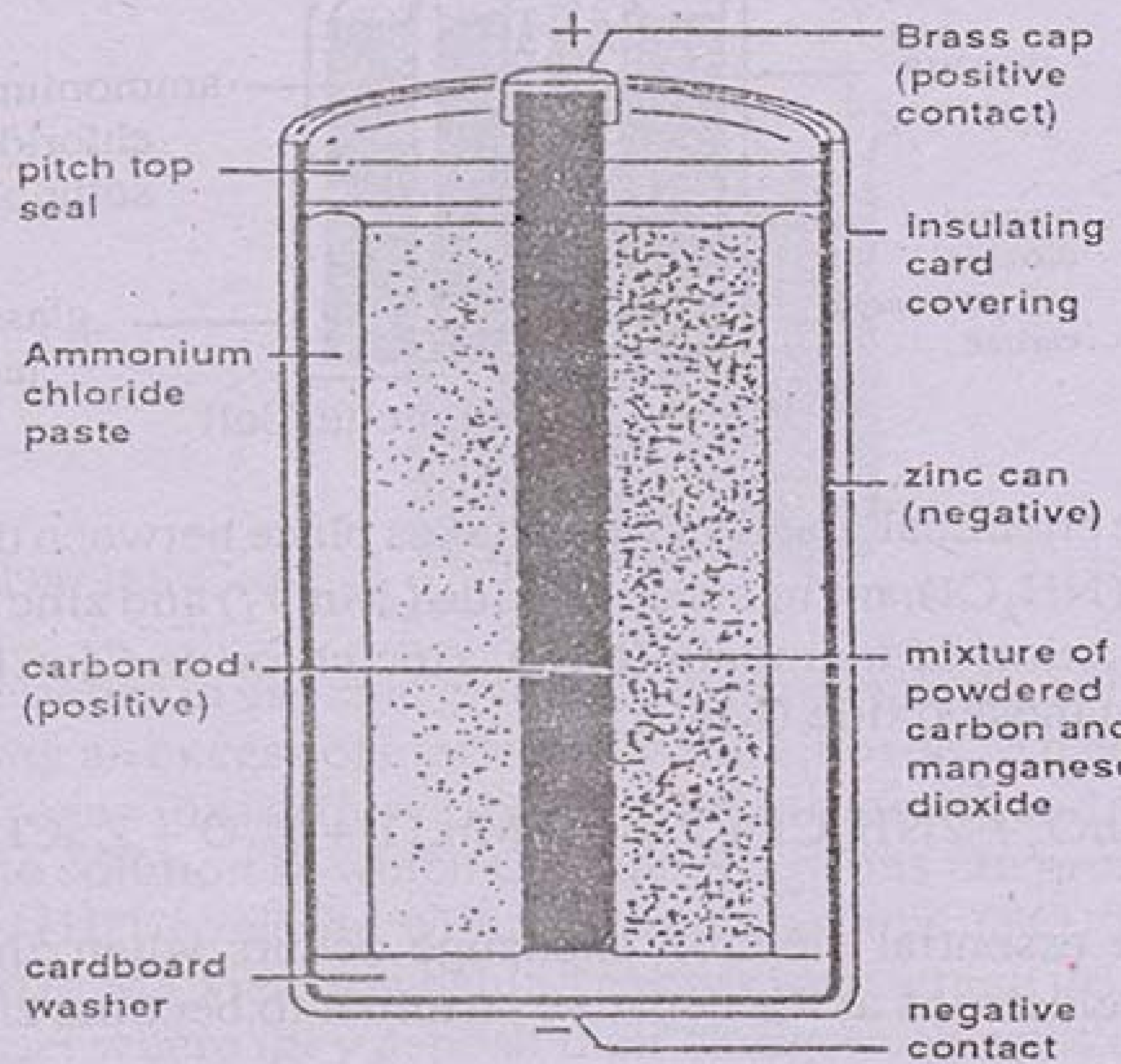


Fig. 16.9 Dry cell

Because of their advantage of being dry and at least leak proof these cells are widely used to supply current in torches, radio sets and transistor appliances.

Secondary Cell or Accumulator

In primary cells like Leclanche's cell, dry cell etc, current is produced by the chemical action between its various components. When the chemicals are exhaused they can not furnish any more

current. The cell has got to be prepared a fresh. There is another class of cells which can be recharged after they have run down by passing a current through them from an external source. This is known as the charging of cell. This is usually carried out from the mains supply. As a result of charging, the cell regains chemical potential energy which is again used to deliver current like a primary cell. This is why such a cell is sometimes known as a storage cell or an accumulator.

Lead-acid Accumulator

One form of a secondary cell in common use is the lead acid accumulator. This accumulator was first devised in 1859 by Pante. Consider the arrangement consisting of a pair of lead plates A and B immersed in a vessel containing dilute sulphuric acid (Fig. 16.10.a). A direct current of about 2 A is passed through the lead plates for a certain period of time, plate A (through which the current enters) will change to a dark brown colour due to chemical reaction with(PbO_2).Now if a voltmeter is connected between A and B it will show that A is 2.2V approximately and positive with respect to B. The arrangement can now function as a cell. Plate A serves as a positive terminal and plate B as a negative terminal. The passage of current through the circuit is from A to B. After some time there will be no passage of current through the circuit and a white layer of lead sulphate ($PbSO_4$) will be formed on both plates. By passing a direct current again through the cell, the lead sulphate on plates A and B can be converted to lead oxide and spongy lead respectively. Hence the cell can once again be used as source of current. This process is called charging the accumulator. The reverse process in which the current is driven out to an external circuit in the reverse direction is known as discharging the accumulator. By charging and discharging the accumulator the thickness of the lead diode layer on plate A gradually increases. As a result the accumulator can be made to supply current for longer periods. In accumulators several lead plates are held parallel to each other in the acid and then connected to the positive and negative electrodes of the cell (Fig. 16.10.a). In modern accumulators instead of lead plate, lead grids are used as shown in Fig. 16.10.b. In the spaces of the grids a mixture of lead oxide (Pb O) and sulphuric acid ($H_2\ SO_4$) is used for the anode plate.

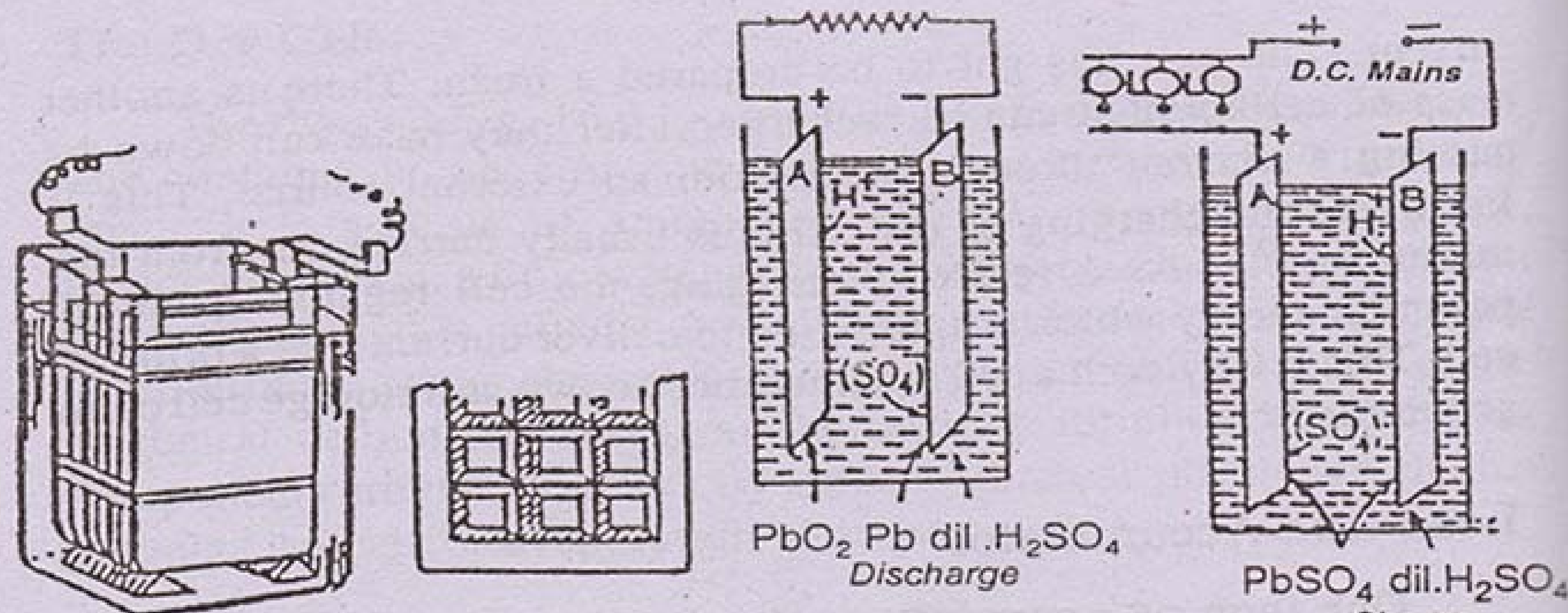


Fig.. 16.10 Lead-acid accumulator

When the cell is working i.e supplying current, the oxide coated plate A from the positive terminal and the metal plate B the negative terminal (Fig.16.10b). When the terminals are connected, by a wire, current flows from A to B through the external circuit (wire) and from B to A through the acid. This action of the accumulator is called discharging. The following chemical action takes place during discharging.

$$PbO_2 + H_2 \longrightarrow PbO + H_2O$$
$$PbO + H_2SO_4 \longrightarrow PbSO_4 + H_2O$$

In the electrolyte sulphate ions migrate to the lead plate and hydrogen ions to the oxide coated plate. On deposition, hydrogen ions change lead peroxide to lead monoxide which forms lead sulphate as shown above.

The lead sulphate is insoluble in the electrolyte and it forms a coating on the plate At the negative plate, (SO_4^{--}) ions also form lead sulphate according to the equation

$$Pb^{++} + SO_4^{--} \longrightarrow PbSO_4$$

when both the plate s are coated with lead sulphate the accumulator is completely discharged and no current is drawn through it.

In order to restore the cell to its original state a current is passed through the cell in the opposite direction. This phenomenon is known as charging of the accumulator and can be achieved by applying a

potential difference from D.C mains so that the positive plate A becomes the anode and the negative plate B, the cathode with respect to the D.C source (Fig. 16.10.c). The D.C source forces a current through the cell in the opposite direction i.e from A to B.

During the charging process H^+ ions migrate to the negative plate B and convert the lead sulphate formed during discharge of the accumulator to metallic lead

$$Pb\,SO_4 + 2H^+ + 2e^- = H_2SO_4 + Pb$$

At the positive plate A, the ions SO_4^{--} produce lead peroxide (PbO_2) and sulphuric acid.

$$Pb\,SO_4 + SO_4 + 2H_2O \rightarrow PbO_2 + 2H_2SO_4 + 2e$$

The net result is the formation of a layer of lead peroxide on the plate A, a layer of spongy lead on the plate B and an increase in the concentration of sulphuric acid. With the removal of $PbSO_4$ layers from the plates of the accumulator it starts supplying current to the external circuit and the accumulator is said to be charged again. Accumulators are widely used in ships, trams, cars and in petrol engines as well as in laboratories.

16.10 ELECTRIC CURRENT

The electrical energy can be easily transferred from one point to another through certain materials known as conductors. A conducting wire such as a copper wire, consists of a large number of free electrons. If the ends of the wire are connected to the terminals of a battery an electric field will be set up within the wire. This electric field will cause a continuous motion of the electrons will be reversed. We therefore, speak of the sources of electrical energy as having fixed polarity. One terminal is called positive and the other negative. The path through which an electric current passes is called an electric circuit.

Earlier scientists had no knowledge of the concept of the electron, so they developed the convention, still in use, that the direction of current is from a positively charged body to a negatively charged body. This current is known as conventional current, whereas the actual flow of current is the motion of electrons from the negative terminal to the positive terminal.

As far as the external effects are concerned it makes no difference if we consider positive charges moving in one direction or negative charges (electrons) moving in the opposite direction. The two ideas are equivalent:

Unit of Current

The rate of flow of charge through a certain cross-section is called current. If the quantity of charge which flows through a unit cross-section during time t is q then current I, is

$$I = \frac{q}{t} \qquad \text{------------} \qquad (16.5)$$

In SI units, the unit of current is called Ampere. Its symbol is A. The submultiple units of current are milliampere (symbol: mA) and micro-ampere (symbol: μ A).

$$1\,mA = 10^{-3}\,A$$
$$1\,\mu A = 10^{-6}\,A$$

Example 16.4

A current is passing through a wire. The total charge passing through its cross-sectional area is 0.6 C in 10 s. Find the quantity of current passing through this wire.

Solution

As $\qquad I = \frac{q}{t}$

Since $\qquad = 0.6C,\ t = 10s$

Therefore $\qquad I = \frac{0.6C}{10s} = 0.06A = 60\ mA$

Example 16.5

Calculate the amount of current through an electric heater if it takes 1800C of charge to heat a room in 3 min.

Solution

Charge,	q	=	1800C
Time,	t	=	3min = 180s
Current,	I	=	?

As $I = \frac{q}{t}$

$$= \frac{1800C}{180s}$$

$$I = 10A$$

Example 16.6

A light bulb with a current of 0.06 A is left burning for 15 min. How much electric charge passes through the filament of the bulb?

Solution

Current,	I	=	0.06A
Time,	t	=	15 min
		=	$15 \times 60 = 900s$
As charge	q	=	I t
		=	$0.06A \times 900s$
		=	54C

Resistance

As you have already learnt that there are large number of free electrons in a conductor. When a potential difference is produced between the two ends of a conductor wire, the free electrons are drifted from the point of higher potential to a point at lower potential. Before the application of potential at the ends of a wire, the free electrons in the wire are in haphazard motion and colliding with each other constantly. The net result of this random motion of electrons is that the number of electrons passing through a section of the wire toward left is same as that passing through the same section toward right. And hence no charge appears to flow through the wire. As we apply a potential across the ends of a wire these electrons are pushed to move

along a certain direction. This directional motion of electrons is called their drift velocity. Thus two types of motion are supperposed, i.e. one the random motion and the other drift velocity. Due to this drift velocity of electrons current is established in a conductor. Since these electrons collide with each other and the cores of the atoms of the body constantly, they find obstacle in their directional motion. This hindrence in their motion is called the resistance of the conductor. The magnitude of this resistance depends upon the nature of the conductor, its length, area of cross-section and the temperature.

Ohm's Law

We are now familiar with the fact that whenever a potential difference (V) is applied across the ends of a conductor, a current (I) starts passing through it. If the value of V is altered then the value of current I is also found to change. Now the question arises as to how the current varies with a change in the applied potential difference. The answer to this question was first found experimentally by a German Physicist George Simon Ohm who discovered that "the current passing through a conductor is directly proportional to the potential difference applied across its ends provided, the temeperature and other physical condition of the conductor are kept constant". This statement is referred to as Ohms law. Mathematically it is expressed as

$$V \propto I$$

$$V = IR \qquad \text{-------------} \qquad (16.6)$$

where R is a constant of proportionality and is known as the resistance of the conductor. It should be noted that R is a physical property of a conductor.

The potential difference-current curve for a given wire at a fixed temperature is a straight line (Fig. 16.7).

Eq. 16.6 can be used to define the unit of resistance. If V=1 volt, I=1 ampere then R=1 ohm. It means if the potential difference across the ends of a conductor is one volt and the resulting current passing through it is one ampere then the resistance offerred by the conductor is one ohm denoted by a Greek letter omega (Ω). The common multiples and submultiples of ohm are the mega ohms (M Ω), kilo ohm (k Ω), milli ohm (m Ω) and micro ohm (μ Ω) which are given below,

1MΩ	=	10^6	Ω
IkΩ	=	10^3	Ω
ImΩ	=	10^{-3}	Ω
Iμ Ω	=	10^{-6}	Ω

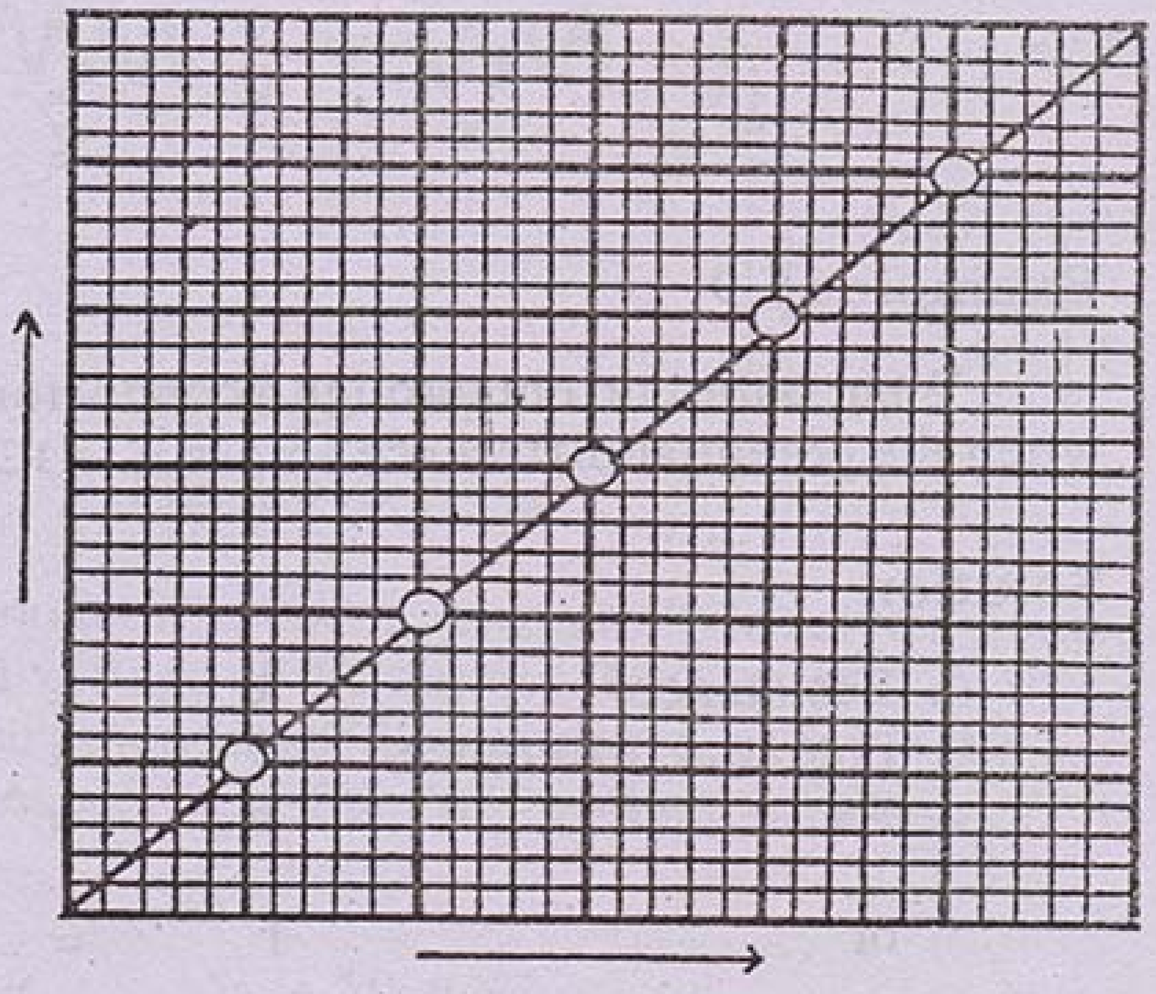


Fig 16.11

Example 16.7

Find the resistance of an electric bulb if 0.60 A current is passing through it and the potential difference across the bulb is 90V.

Solution

Current,	I	=	0.60A
Potential difference,	V	=	90V
Using Ohm's Law	V	=	IR
or	R	=	$\frac{V}{I} = \frac{90\text{V}}{0.60\text{A}}$
	R	=	$1.5 \times 10^2\ \Omega$

Example 16.8

What is the potential difference across an electric iron of resistance 27.4 Ω when the current through it is 8.75 A?

Solution

$$\text{Resistance,} \quad R = 27.4\,\Omega$$
$$\text{Current,} \quad I = 8.75A$$

Now using Ohm's law

$$V = IR$$
$$V = (8.75A)\,(27.4\,\Omega)$$
$$= 239.75V = 240V$$

Example 16.9

What is the current through a conductor with a resistance of 19 Ω when the potential difference across it is 120V?

Solution

$$\text{Resistance,} \quad R = 19\,\Omega$$
$$\text{Potential difference} \quad V = 120V$$
$$\text{As} \quad V = IR$$
$$\text{or} \quad I = \frac{V}{R} = \frac{120V}{19\Omega}$$
$$I = 6.3\text{ A}$$

16.11 ELECTRIC CIRCUIT AND COMBINATION OF RESISTANCES

In the section 16.10 on electric current we have discussed the most simple electrical circuits in which a single source of emf and external resistances are connected. In actual practice circuits may comprise of number of circuit components (resistors, capacitors, bulbs, motors, etc), and a number of sources of emf. We shall use the term resistance or resistor for a circuit element to represent them in a circuit. In electrical circuits the resistors are interconnected in two ways (i) Series (ii) Parallel.

(i) Resistances in Series

In such a combination the resistors provide a single path to the passage of current and the same current passes through each of the

resistors. In (Fig. 16.12), resistors R_1, R_2, and R_3 are joined together in such a way that the same current I passes through each of these resistors. These resistances R_1, R_2, and R_3 are in series. Series combination of resistances has the following characteristics.

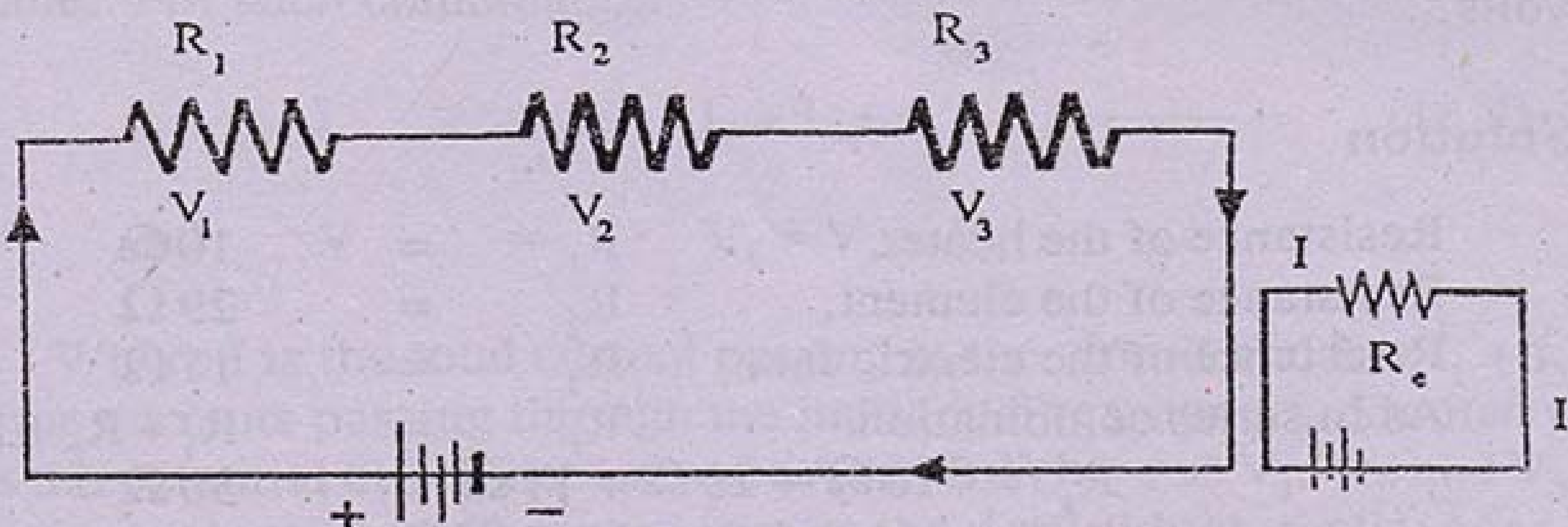


Fig. 16.12

(i) The sum of potential differences across individual resistors is equal to the total voltage of the battery. In (Fig. 16.12), if V_1, V_2, and V_3 are the potential differences across the resistances R_1, R_2, and R_3 respectively then the voltage of battery V is given by,

$$V = V_1 + V_2 + V_3$$

(ii) The resistances R_1, R_2, and R_3 can be replaced by a single resistance R_e which when joined in the circuit in place of these resistors allows the same current to pass through it as can pass in the presence of the individual resistors. This combined or equivalent resistance (R) of the series circuit can be obtained by adding individual resistances combined in series. Since

$$V = V_1 + V_2 + V_3$$

$$I R_e = I(R_1 + R_2 + R_3)$$

$$R_e = R_1 + R_2 + R_3 \quad \text{----------} \quad (16.7)$$

If more than one resistances were in series we would continue to add them together to get the total equivalent resistance, i.e. $R_s = R_1 + R_2 + R_3 + R_4 \ldots\ldots\ldots\ldots R_N$ for N resistances in series.

Example 16.10

Find the equivalent resistance of a series circuit containing $10\,\Omega$ heater, $29\,\Omega$ element and $11\,\Omega$ electric fan. What will be the current passing through the circuit if it is driven by a voltage source of 100 volts?

Solution

Resistance of the heater,	R_1	=	$10\,\Omega$
Resistance of the element,	R_2	=	$29\,\Omega$
Resistance of the electric fan,	R_3	=	$11\,\Omega$
As in series combination,	R_s	=	$R_1 + R_2 + R_3$

$$R_s = 10\,\Omega + 29\,\Omega + 11\,\Omega = 50\,\Omega$$

and the potential difference, $V = 100V$

By using ohm's law, $V = IR$

or

$$I = \frac{V}{R} = \frac{100V}{50\Omega} = 2A$$

(ii) Resistances in Parallel

When two or more resistances are joined in such a way that one end of each resistor is connected to one terminal of the battery while the other ends are connected to the second terminal of the battery the resultant combination is known as a parallel combination of resistors. This is shown in (Fig. 16.13). When resistances are connected in parallel different amounts of current pass through each resistance so that the sum of all such currents is equal to the total current supplied by the battery.

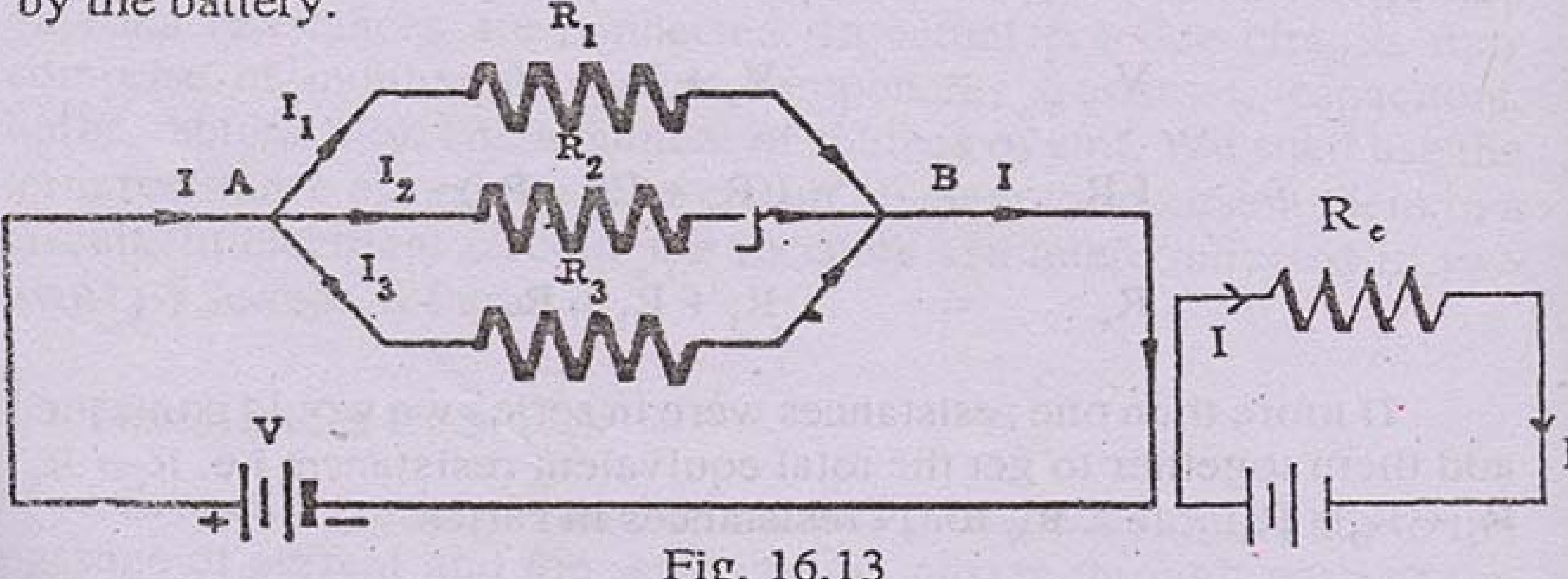


Fig. 16.13

However, the potential difference across each resistor is the same and is always equal to the voltage of the source. In (Fig. 16.13), the current branches off into three paths at the point A and recombines at B. Thus we can say that the resistors R_1, R_2 and R_3 have been connected in parallel. For such combination

$$I = I_1 + I_2 + I_3 \quad \text{---------} \quad (16.8)$$

and

$$V = V_1 = V_2 = V_3$$

Where I is the total current passing in the circuit and $I_1 + I_2 + I_3$ are the currents passing through the individual resistances. Similarly V is the potential difference between A and B. Where as V_1, V_2, and V_3 are the potential difference across the individual resistances. Substituting the values of I_1, I_2 and I_3 in terms of V_1, V_2, and V_3 in the Eq. 16.8, we get

$$\frac{V}{R} = \frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3}$$

As

$$V = V_1 = V_2 = V_3$$

So,

$$\frac{V}{R} = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3}$$

Eliminating V the equation becomes

$$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$$

This equation shows that the resultant resistance of the resistors joined in parallel is less than the least resistance of the component resistors (Fig.16.13). If more than 3 resistances are connected in parallel then the total equivalent resistance R_e of N resistances is given by

$$\frac{1}{R_e} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \ldots\ldots \frac{1}{R_n} \quad \text{---------} \quad (16.9)$$

Example 16.11

Find the equivalent resistance when a $4\,\Omega$ bulb and an $8\,\Omega$ resistor are connected in parallel.

Solution

Resistance of the bulb, $R_1 = 4\,\Omega$

Resistance of the resistor, $R_2 = 8\,\Omega$

For parallel combination, $\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2}$

$$R = \frac{R_1 R_2}{R_1 + R_2}$$

$$R = \frac{4\Omega \times 8\Omega}{4\Omega + 8\Omega} = \frac{32\Omega^2}{12\Omega} = 7.2$$

Example 16.12

What resistance would have to be added in parallel with a $80\,\Omega$ electric hair dryer to reduce the equivalent resistance to $16\,\Omega$.?

Solution

Resistance of hair dryer, $R_1 = 80\,\Omega$

Resistance to be added, $R_2 = ?$

Equivalent resistance, $R = 16\,\Omega$

As $\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2}$

$$\frac{1}{R_2} = \frac{1}{R} - \frac{1}{R_1}$$

or $R_2 = \frac{RR_1}{R_1 - R}$

$$R_2 = \frac{16\Omega \times 80\Omega}{(80-16)\Omega}$$

$$= \frac{1280\Omega^2}{64\Omega}$$

$$R_2 = 20\,\Omega$$

Example 16.13

An incandescent bulb, when hot, has a resistance of 60 Ω while a hot lamp has a resistance of 120 Ω. Calculate the current through each bulb, the total current in the circuit and the total circuit resistance when both bulbs are connected in parallel and operate from a 60V supply.

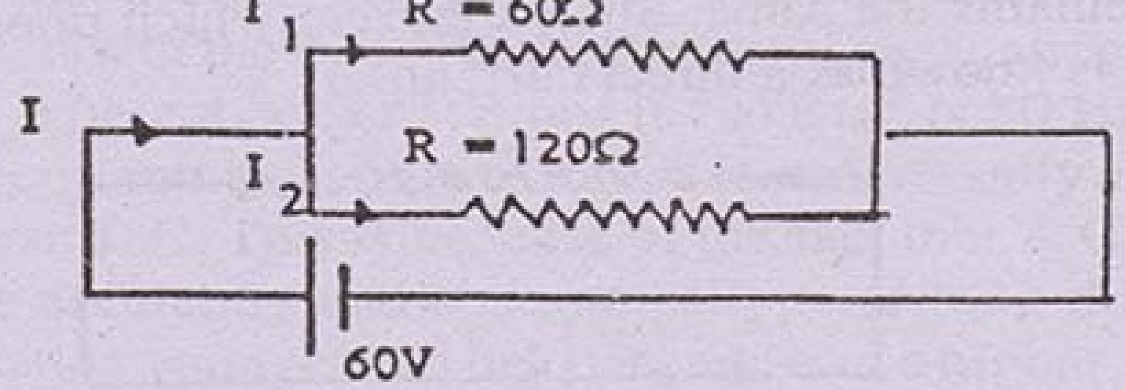


Fig. 16.14 Lamps in parallel

Solution

Resistance of the incandescent bulb	R_1	=	60 Ω
Resistance of the hot lamp,	R_2	=	120 Ω
Potential difference,	V	=	60V

Current through incandescent bulb

As $\quad V = I_1R_1$

or $\quad I_1 = \dfrac{V}{R_1} = \dfrac{60V}{60\Omega} = 1A$

Current through the hot bulb

$$I_2 = \frac{V}{R_2} = \frac{60V}{120\Omega} = 0.5A$$

Total current I_t will be equal to the sum of individual current

$$I_t = I_1 + I_2$$

$$I_t = 1A + 0.5A = 1.5A$$

Total circuit resistance by Ohms law is given as

$$R = \frac{V}{I} = \frac{60\ V}{1.5\ A} = 40\Omega$$

16.12 DIRECT CURRENT AND ALTERNATING CURRENT

D.C. stands for Direct Current. Such a current can be obtained by connecting the two ends of a conductor to the terminals of a battery as shown in Fig. 16.15. A current will start to pass from the positive to the negative terminal of the battery in the external circuit and its direction always remains the same. Such a current which does not change its direction is known as a direct current.

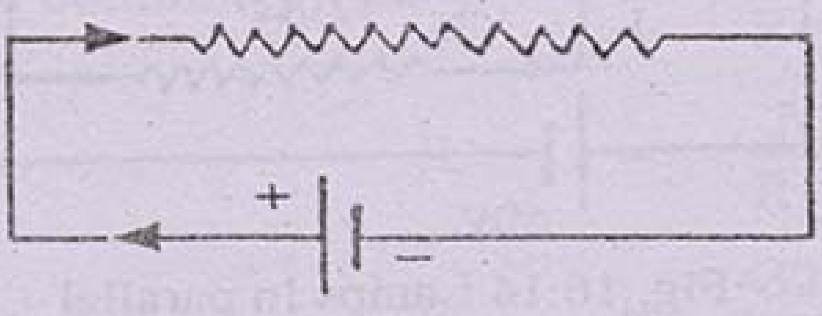


Fig. 16.15 Current through a conductor

Thus direct current is like water flowing out from a tap to a garden hose. This similarity is due to their unidirectional motion. But if a current changes its direction many times a second it is known as alternating current or A.C. Such current can be obtained if a resistor is connected in series with a source of alternating voltage as shown in Fig. 16.16.a.

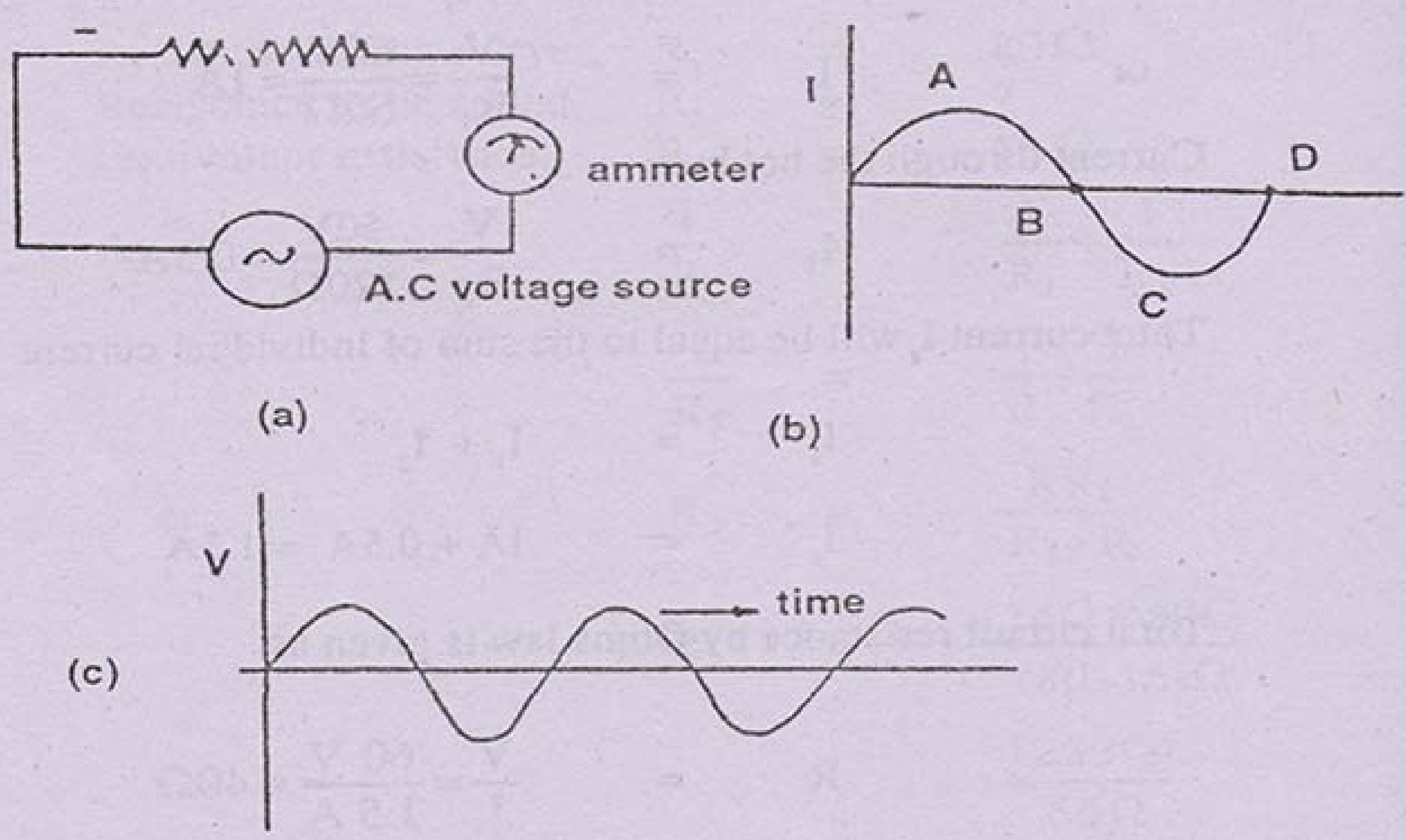


Fig. 16.16 Direct current and alternating current

The current passing through the external circuit as indicated by the A.C. meter is of the form shown in Fig. 16.16.b. Since the current changes its direction alternatively from the positive half to the negative half such a current is termed as Alternating Current. The time taken by the current to complete one cycle is called the time period (T).

When the A.C. source supplies continuous energy to the external circuit a continuous alternating current will be obtained as shown in Fig. 16.16.c. The number of cycles of current completed in one second is called the frequency of the A.C. The frequency of alternating current generated by the different power stations in Pakistan is 50 Hz.

Alternating currents and voltages have been universally accepted as more useful than D.C. This is because of the fact that A.C. can be transmitted more efficiently and economically over long distances. Moreover, alternating voltages can be changed from a higher value to a lower value and vice versa with the help of a transformer. It can also be converted into D.C. easily for use in those industries which require direct currents e.g., in electroplating.

16.13 ELECTRICAL ENERGY AND JOULE'S LAW

There are numerous application of the heating effect of an electric current. Can you imagine life without electric cookers, electric kettles, electric fires, hair dryers, soldering irons etc? All of these work as they do because when electric charge passes along a wire, the wire gets hot.

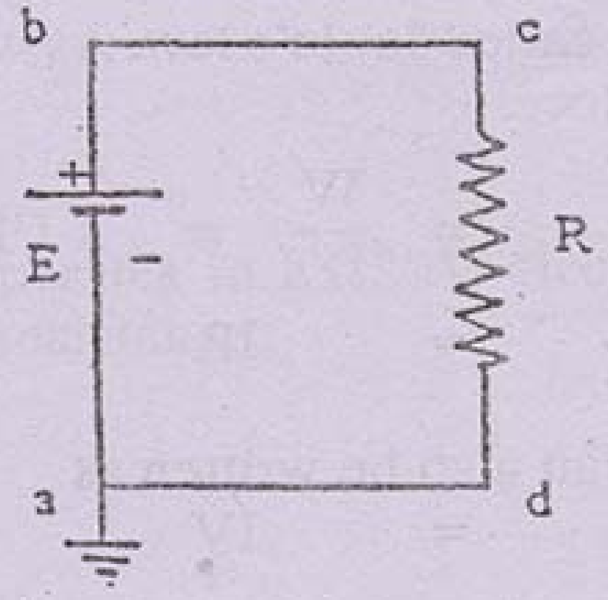


Fig. 16.17 A circuit consisting of a battery and resistance R

If a battery is used to establish an electric current in a conductor there is á continuous transformation of chemical energy into other forms of energy (Fig.16.17). As you learn that when a body fall from

a height, the potential energy decreases and the energy lost is changed into some other from of energy. The same applies in electricity when charge flows from a higher to a lower potential level, energy is released. If you imagine that the current in a circuit is formed by 'drops' of electricity, each having a charge of I coulumb and carrying equal sized 'bundles' of elecrical energy. As a 'drop' moves around the circuit it gives up all its energy which is changed to other forms of energy. Note that electrical energy is 'used up' not charge or current. If one coulumb of charge moves through a potential difference of one volt in the direction of the electric field then one joule of energy is released. Hence, when charge flows, energy W in joule is, equal to the product of potential difference (V) and charge (q).

i.e. $W = q \times V$

as $I = \dfrac{q}{t}$

so $= I \times t,$

and $W = I \times t \times V$

From ohm's law,

$$V = IR$$

then $W = I \times t \times RI = I^2tR$

I^2Rt is the amount of electrical energy dissipated as heat energy. This relationship is called Joule's law. As we know that power is work per unit time therefore.

$$P = \frac{W}{t} = I^2R \qquad \text{----------} \qquad (16.9)$$

Since $V = IR$

Then power can also be written as

$$P = IV$$

The unit of power is the watt and the unit of time is the second hence the unit of energy is the watt second. But this is a small unit and for normal use the time is most conveniently taken in hours so that unit of electrical energy is taken as kilo Watt hours (kwh). Thus

Energy in kWh $= P \times t \times 1000$

where P is expressed in watts and 't' in hours. The kilo watt hour is the commercial unit in which the electrical energy supplied by WAPDA is measured and bills prepared.

Example 16.14

An electric heater has a resistance of $20\,\Omega$ and it works when a potential difference of 220V is applied across its terminals. Find the current passing through the heater and the power rating of the heater.

Solution

Resistance of the heater,	R	=	$20\,\Omega$
Potential difference,	V	=	$220V$
As	V	=	IR
	I	=	$\dfrac{V}{R}$
or	I	=	$\dfrac{220\ V}{20\ \Omega} = 11A$
Also as we know that	P	=	I^2R
			$= (11A)^2 \times (20\,\Omega) = 2420W$
		=	$2.420\ kW$

Example 16.15

A 100 watt bulb operates in a 220V circuit. Find the current through the bulb and its resistance.

Solution

Power,	P	=	$100W$
Potential difference,	V	=	$220V$
As	P	=	IV
or	I	=	P/V
then	I	=	$\dfrac{100\ W}{220\ V} = 0.45\ A$

Using Ohm's law, $V = IR$

or $R = V/I = \frac{220\ V}{0.45\ A} = 488.8\Omega$

16.14 HOUSE CIRCUITS

Electricity usually comes to our homes by two wires or lines, the live (L) and the neutral (N). The potential difference between these wires is 220V. The neutral wire is connected to the earth about every 100 m. The potential on the live wire is alternately positive and negative with respect to the neutral wire. Study the house circuit as shown in Fig. 16.18.

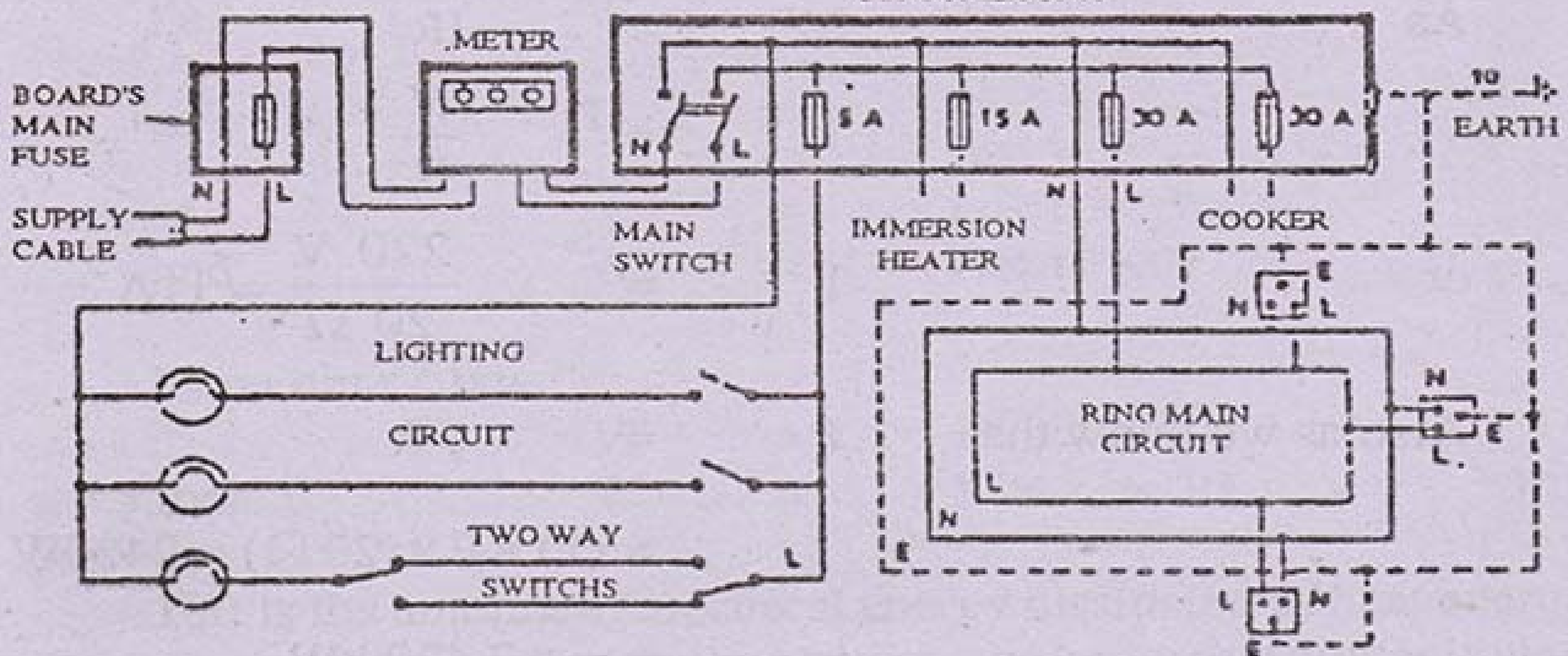


Fig 16.18 A typical household circuit

(a) Circuits in parallel

Every circuit is connected in parallel with the supply i.e across the live wire and neutral wire and receives the full mains p.d of 220V.

(b) Switches and fuse

These are always in the live wire side of the circuit. If they were in the neutral side, then lamp and power sockets would be 'live' even when switches were 'off' or fuses 'blown'. A shock (fatal) could then be obtained by a person who touches the element of an electric fire when it was switched off.

(c) Stair care circuit

The lamp is controlled from two places by two way switches.

(d) Ring main circuit

The live and neutral wires are thick (called heavy duty wires) and each run in a complete ring around the house; and the power sockets, each rated at 13A, are tapped off from them. Thinner wires can be used for each socket since the current to each socket is less than 13 A. In the whole ring the ring has a 30A fuse and if it has ten sockets, all can be used so long as the total current does not exceed 30A.

(e) Fused plug

Only one type of plug is used in a ring main circuit. It is wired and has its own cartridge fuse, 3A (blue) for appliances with powers upto 660 W and 13A (brown) for those between 720 W and 2860 W

(f) Earthing and safety

A ring main circuit has a third wire which is connected to one part (E) of the socket. This wire is earthed by being connected to a metal water pipe in the house or to an earth connection on the supply cable. We take the earth itself as being at zero potential. This is purely a matter of convenience. It does not mean that the earth has no electrical charge. Actually the earth has a negative charge but this is so large that any charge given to or taken from it has a negligible effect.

This third wire is a safety precaution to prevent an electric shock should an appliance develop a fault. The earth pin (E) in a plug is connected to the earth connection through a path having almost zero resistance. If the element of an electric fire breaks or sags and touches the case a large current flows to the earth and blows the fuse. Otherwise the case becomes live and anyone touching it will receive a shock which might be fatal, especially if they were earthed by standing in a damp environment, e.g on a wet concrete floor. Circuit breakers are also used instead of fuses for safety. The circuit breakers are connected in series with the wire entering the house. The circuit breaker is a device which break the circuit when the current exceeds the rated value (5A to 13A). These are made on the principle of a electromagnet which will be discussed in chapter 17.

SUMMARY

— There are two categories of material objects, one through which current can pass and the second through which current can not pass. Objects belonging to the first category are copper, aluminium, silver, iron or, in general, most metallic objects called conductors. The second category of materials through which electric current can not pass so easily are called insulators. Plastic, rubber, nylon, cotton, wood etc are typical examples of insulators. Conductors have a large amount of free electrons where as insulators have an insignificant amount of these electrons.

— There is, however, an intermediate category of materials called semiconductors.

— According to Coulomb's law, the force of attraction and repulsion between two charges q_1 and q_2 separated by a distance r, is given by

$$F = K\frac{q_1q_2}{r^2}$$

— Electrostatic Induction is the process in which a charged object charges another neutral object without touching it.

— The electric field due to a charge can be defined as the space around that charge. In which its influence would be felt by another charge placed there.

— The electric field strength or electric field intensity $E = F/q_o$ i.e, the force per unit charge.

— Electric potential determines the flow of charge. The potential difference between two points A and B may be defined as the work done in carrying a unit positive charge from A to B without imparting any acceleration to the charge. The unit of potential difference is the volt. The potential difference between two points will be one volt if one joule of work is required to move a positive charge of one coulomb from one point to another.

— Electromotive force (e.m.f) of a source of electrical energy such as a cell is the work done in driving charge that passes through the source.

— The difference between PD and emf is that a source of e.m.f in a circuit does work on the moving charges (supplies energy) where as P.D refers to the work done by the charges (energy expended) in passing through the resistance.

— Electric current may be defined as the rate of flow of charge.

— Conventional electric current is taken as an electric fluid which flows from the positive terminal of the source of emf to the negative terminal.
— The unit of current is called ampere. One ampere of current is said to be flowing in a conductor if one coulomb of charge flows across any cross section of the wire in one second.
— Resistance is an opposition to the motion of electrons. This arises as a result of collisions of moving electrons with the neighbouring atoms.
— The current passing through a conductor is directly proportional to the potential difference applied across the ends of a conductor provided that the temperature and other physical conditions of the conductor do not change. This is called ohm's law.
— The ohm is the unit of resistance. If a potential difference of 1 volt is applied across the ends of a conductor and the resulting current flowing through the conductor is one ampere then the resistance of the conductor is one ohm, denoted by the Greek letter omega (Ω).
— The path through which an electric current passes is called an electric circuit. If there is a break in the circuit no current will pass through the circuit. Such a circuit is called an open circuit.
— Resistances are said to be connected in series if the same current passes through all the resistors and the resistors provide a simple path to the flow of current. if resistor R_1, R_2, R_3 ----- R_n are connected in series then the resultant equivalent resistance R_e is given by,

$$R_e = R_1 + R_2 + R_3 + \text{---------} R_n$$

Resistances are said to be connected in parallel if two or more resistors are joined in such a way that one end of each resistor is connected to one terminal of the battery while the other ends are connected to the second terminal of the battery. For a number of resistances R_1, R_2, R_3 —— R_n to be connected in parallel the resultant resistance R_e is given by,

$$\frac{1}{R_e} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \frac{1}{R_4}$$

— A battery is a source of electrical energy having fixed polarity and terminals.
— A simple cell has a + ve copper and -ve zinc plate and an electrolyte of dilute śulphuric acid. It can supply continuous electric current for some time which then diminishes due to its defects known as polarization and local action. Polarization is due to the generation of hydrogen

at the copper plate which insulates it from the electrolyte and hence stops the flow of current. In local action, the plate is being dissolved in the electrolyte due to local currents.

— A leclanche dry cell has a +ve carbon and a - ve zinc plates, an electrolyte of ammonium chloride paste and a solid depolarizer of manganese dioxide mixed with powered carbon. The hydrogen produced at the carbon plate is oxidised to water. The emf of this cell is about 1.5 volts and the current is only maintained for a short time. Both simple and leclcanche cell can not be recharged.

— Direct current (D.C.) can be made to flow through a conductor if it is connected to the terminals of a battery. The direction of this current is always from the positive to the negative terminal of the battery i.e., a direct current always flows in one direction.

— Alternating current changes its direction many times a second. Such a current can be obtained if a resistor is connected in series with a source of alternating voltage.

— The Energy consumed in an electrical circuit is given by $P = I^2Rt$. This is known as Joule's law.

QUESTIONS

16.1 Write answers to the questions given below:

(i) State Coulomb's law and define the unit of charge.

(ii) What do you understand by an electric field? Define the intensity of electric field.

(iii) Define electric potential difference between two points of an electric field.

(iv) What do you understand by a capacitor and its capacitance. Define its unit.

(v) Define electric current and write its equation.

(vi) What do you mean by Conventional current?

(vii) State and explain Ohm's law.

(viii) Explain series and parallel combination of resistances.

(ix) If a current flows through a resistance heat is dissipated. On what factors, does this heat depend.

(x) What is meant by polarization?

(xi) What do you mean by depolarization?

(xii) Distinguish between A.C and D.C.

(xiii) Write short notes on:
(a) Leclanche Cell (b) Daniell Cell (c) Dry Cell.

16.2 Fill in the blanks.

(i) Practically the whole _______ of the atom is concentrated in its nucleus.

(ii) The electrons of the outer most orbit in an atom are known as ________ electrons.

(iii) The smallest unit of negative charge that can exist in nature is on an ________.

(iv) The force of attraction or repulsion between two charged bodies is given by
F = ________.

(v) Coulomb is defined as the amount of charge carried by a current of ________ in _______ second.

(vi) Like charges ________ each other and unlike charge ________ each other.

(vii) The number of ________ electrons are very _________ in a insulator.

(viii) One milliampere is equal to _________.

(ix) The graph between current passing through a conductor and the potential difference across its ends is always a ________.

(x) If R_1, R_2, R_3 resistances are connected in parallel, the resultant resistance R is given by I/R = __________.

(xi) If resistance R_1 and R_2 are connected in such a way that potential difference across them always remains the same, this will be possible only when they are connected in ________.

(xii) The rate of flow of electric charge in a conductor is called ________.

(xiii) In the S.I. system of units electric field intensity is measured in _________ per _________.

(xiv) According to Ohms law, the ________ flowing in a conductor is ________ proportional to the potential difference between its ends.

(xv) In a parallel combination of resistors the total resistance is always ________ than the resistance of any individual resistor.

(xvi) The potential difference between two points is 1 ________ if 1 joule of electrical energy is changed into another form of energy when ________ passes from one point to the other.

16.3 Given below are a few possible answers to each statement; Identify the correct one.

(i) _______ revolve around the nucleus in their respective orbits.
(a) Neutrons (b) Protons (c) Electrons

(ii) The lightest particle in an atom is _______
(a) neutron (b) proton (c) deutron (d) electron

(iii) If the quantity of charge on each of the two bodies is doubled, the force between them becomes ________
(a) twice (b) four times (c) nine times (d) sixteen times

(iv) The value of constant that occurs in coulombs force formula is ________ Nm^2/C^2
(a) 9×10^{-9} (b) 9.0×10^{-16} (c) 9.0×10^{9} (d) 9.9×10^{-9}

(v) If two joules of energy is required to transfer one coulomb of charge from one point to another, the potential difference between these points will be ________ volt.
(a) one (b) two (c) four (d) six

(vi) The substance used as a medium between the two plates of a capacitor is knwon as ________
(a) conductor (b) semi-conductor (c) dielectric (d) electrolyte

(vii) If the charge on a plate of a capacitor is doubled, the potential difference between its two plates will become ________
(a) half (b) double (c) four times (d) one third

(viii) The unit of capacity is known as ________
(a) coulomb (b) volt (c) ampere (d) ohm (e) farad

(ix) ________ is that which is connected in series with the line wire in the electric circuit of a house.
(a) Galvanometer, (b) Voltmeter, (c) Fuse, (d) Ammeter.

(x) The commercial unit of electric energy is known as ________.
(a) ohm, (b) volt, (c) ampere × second,
(d) coulomb, (e) watt-hour, (f) kilo-watt-hour.

(xi) Electric power in watts is obtained by the product of ________
(a) volt and coulomb, (b) current and resistance,
(c) coulomb and ampere, (d) volt and ampere.

(xii) The maximum value of the alternating current in either direction is known as its ________.
(a) average, (b) square, (c) root mean square, (d) peak value

(xiii) One mega ohm resistance is equal to ________ ohm.
(a) 10^6, (b) 10^{-6}, (c) 10^8, (d) 10^2, (e) 10^{-8}

(xiv) In a gas discharge tube the electric current is due to flow of __________.

(a) electrons, (b) positive ions, (c) neutrons, (d) electrons, (e) positive ions.

(xv) which one of the following is equivalent to "joule per coulomb".

(a) ampere, (b) ohm, (c) volt, (d) watt.

16.4 Pick out true and false from the following sentences.

(i) The mass of a proton in 1000 times larger than the mass of an electron.

(ii) Neutron is a particle which has a negative charge and its mass is equal to that of an electron.

(iii) The force between two charges varies directly as the square of the distance between them.

(iv) Repulsion is the sure test for a body being charged.

(v) Objects build up a static charge when protons are lost from it due to friction.

(vi) Electric intensity at any point in an electric field is the force experienced by an electron placed at that point.

(vii) Generation of negative charges in a neutral body due to presence of a charged body near it is known as electrostatic induction.

(viii) The potential difference between two points in an electric field is the force exerted on a unit positive charge as it is moved between these two points.

(ix) The electrons in metals which are not attached to any particular atom are known as free electrons.

(x) The flow of current in a metal is due to flow of protons.

(xi) Volt is the unit of electric current.

(xii) The conventional current always flows from a point at a higher potential to a point at lower potential.

(xiii) The current passing through a conductor is proportional to the square of the potential difference across its two ends.

(xiv) When the resistances are connected in such a way that a number of paths are provided to flow of electric current, the resistance are said to be connected in series.

(xv) Alternating current is that which changes its direction of flow after a definite period of time.

(xvi) The unit of electric power is ampere.

PROBLEMS

16.1 Calculate the amount of work done in carrying a charge of $+2.5\,\mu C$ from A to B if A is at a potential of - 60V and B is at + 10V.

(1.75×10^{-4})

16.2 Find the potential difference between the two ends of a conductor if it offers a resistance of $5\,\Omega$. Take the current flowing through the conductor as 5 amperes.

(25 Volt)

16.3 The potential difference applied to the terminals of a portable radio is 9.0 Volts. Find the resistance of the radio if a current of 25 mA is flowing through it.

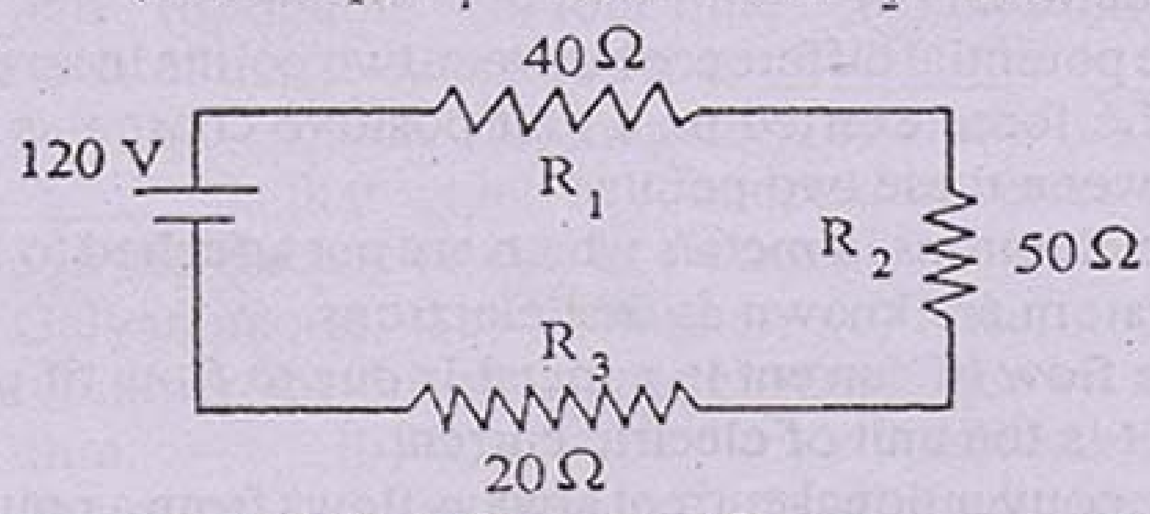
($360\,\Omega$)

16.4 An electric toaster has a resistance of $12\,\Omega$. What current will it draw from a 120 V supply?

(10 amp)

16.5 A series circuit consisting of three resistors having resistances of $40\,\Omega$, $50\,\Omega$ respectively, is connected across a voltage source of 120V as shown in fig. below. Find the current in the circuit and th potential.difference across each resistor.

(1.09 amp, V_1 = 43.64 Volt, V_2 = 54.5V and 21.8V)



16.6 Resistances of $4\,\Omega$, $6\,\Omega$ and $12\,\Omega$ are connected in parallel and then joined to an emf source of 6V. Find the value of
(i) the equivalent resistance of the circuit
(ii) the total current from the circuit
(iii) the current passing through each resistance

[(i) $2\,\Omega$, (ii) 3 amp (iii) 1.5 amp, 1 amp and .5 amp)]

16.7 Two light bulbs of 100 watt and 60 watt both operate in a 220V circuit. Which bulb has the higher resistance and which bulb carries the greater current?

(60 watt bulb . , 100 watt bulb)

CHAPTER - 17

MAGNETISM AND ELECTROMAGNETISM

17.1 INTRODUCTION

Magnetism is one of the oldest of sciences. It was discovered about three thousand years ago. In some books it is reported that the Chinese were the first to discover magnetite, a magnetised iron ore, but in others it is also mentioned that the Greeks had discovered it first. The story in each case is same, that a wood cutter put his axe on a rock while he was resting in the forest. When he got up to start his work again he found that his axe was attracted by rock. After some more observations it was found that the rock attracted samll objects which were made of iron. This was the first magnet found on the earth.

At present magnets are artificially made from the alloys of iron, cobalt and nickel. These alloys are used because the pure elements can not retain their magnetic properties permanently. All the physical properties which are related to magnets are known as magnetism. At present magnetism is a vast field of Physics and it has many applications. Some of the things that you are familiar with are magnetic tapes, computer diskettes, the electric power generators, electric motors and transformers. In all of the above and many other areas magnetism plays an important role. Some of these will be discussed later in this chapter.

17.2 MAGNETIC POLES

Every magnet has two poles, a North seeking pole and a South seeking pole on its opposite ends. If a magnet is allowed to rotate freely it will come to a rest position with one end facing geographical North and the other end facing geographical South (Fig. 17 .1). If we mark the end facing North with the letter N and disturb the orientation of the magnet it will again come to rest with the end marked N towards the North.

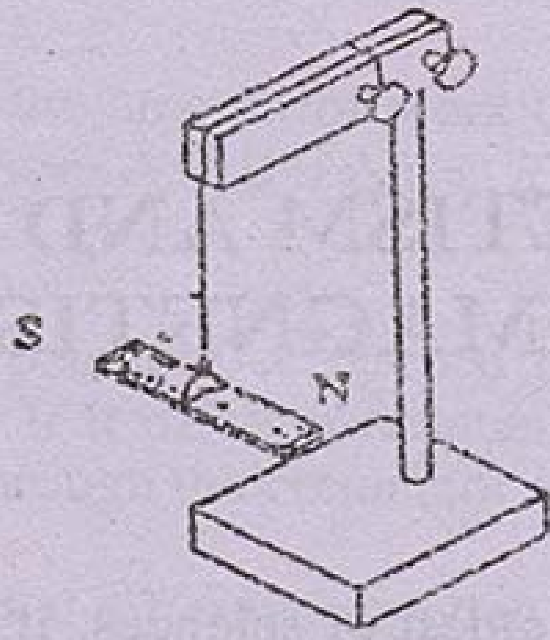


Fig. 17.1 A magnetized piece of iron points N and S if suspended.

When a magnet is dipped in iron filings the filings are attracted to the magnet. There is a greater concentration of the filings at the two ends than at the rest of the body of the magnet. If two magnets are brought near each other the North pole of each attracts the South pole of the other. The two North poles repel each other and the two South poles do the same as well. In other words like poles repel each other and unlike poles attract each other.

The strength of attraction varies with different magnets and will be discussed later but in a magnet every pole is coupled with an opposite pole of the same strength or power as itself.

17.3 MAGNETIC FORCE

It can be shown experimentally that magnets attract other magnetic substances. This attractive force which is different from both the electrostatic and gravitational forces is called Magnetic force. We will learn more about the nature of magnetic force from the following experiment.

Take two bar magnets and identify their North and South poles. Bring the North pole of one magnet closer to the South pole of the other magnet. We will notice that the two magnets attract each other even before touching each other. We may also notice that the force of attraction increases as the distance decreases between the poles . Now if we bring the North pole of the magnet closer to the North pole of the other magnet, as stated at above, the two poles will repel each other with a repulsive magnetic force. If we do the same experiment with the South pole we will see that the South pole is also repelled by a South

pole and attracted by a North pole of the other magnet. From this experiment one can realize the similarity between two types of electric charges and two types of magnetic poles that the like poles have a repulsive magnetic force and opposite poles have an attractive magnetic force.

The magnetic force also resembles in behaviour to a gravitational force between two masses or the electric force between two-electric charges i.e., action at a distance. The magnetic force between two poles depends on the distance between them. Take two magnets and hold them close to each other, you will feel a strong force. On moving them apart from one another the force decreases as the distance increases. A more careful observation will reveal that the force is inversely proportional to the square of the distance between the poles. The force between two poles also depends on the strength of the poles. The force is stronger when a powerful magnet acts on another magnet or a piece of iron.

17.4 MAGNETIC FIELD

The space surrounding a magnet in which its magnetic effect is felt is called a "Magnetic field". It is a region within which the magnet can exert its magnetic force. For example a north pole of a magnetic needle will experience a force in the magnetic field and the needle is deflected. Also an iron piece will be attracted towards the magnet when placed in the magnetic field.

17.5 MAGNETIC FIELD OF THE EARTH

A freely suspended magnet, or a compass, always lines up in a North South direction. If it is disturbed it again comes to the North-South position after a few oscillations. A small compass in the presence of a magnetic field will rotate until its N-pole points in the direction of the field. If there is no magnet present, the needle is influenced only by the earth's magnetic field which causes the needle to point in the North-South direction. This discovery was used by the early navigators to find the direction of North.

The earth behaves as a large bar magnet along the North South direction with the North pole towards the South geographic pole. (Fig. 17.2). The actual cause for the earth's magnetic field is not exactly known. All the theories, which are currently followed, assume that the

earth's magnetic field is similar to that of an imaginary bar magnet situated at its centre. It is quite clear that it has two polarities like an ordinary bar magnet.

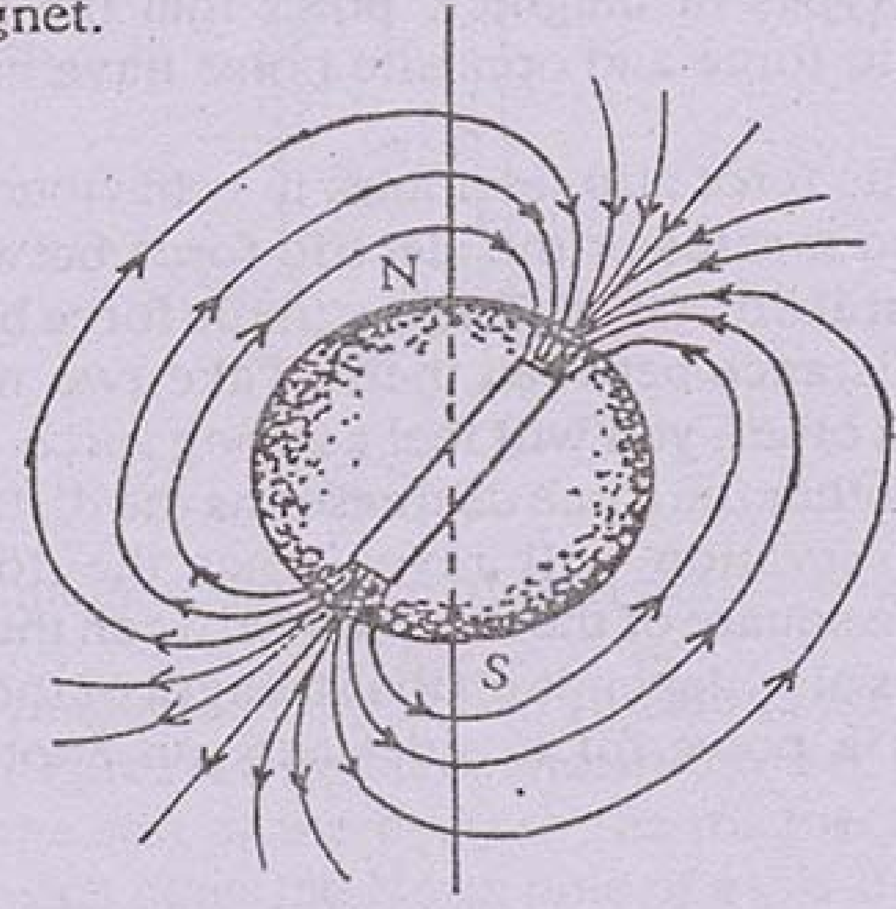


Fig. 17.2 Earth's magnetic field

17.6 MAGNETS AND MAGNETIC MATERIALS

We have already stated that substances which attract each other megnetically are made of metals like iron, nickel and steel. They are called magnets and always point in a particular direction when freely suspended in the air. The substances which are neither attracted nor repelled by a magnet are called non-magnetic substances. We can see that wood, glass, paper etc. are not affected by a magnet.

If we bring nails or paper pins to the pole or the magnet they will stick to the magnet (Fig.17.3a).

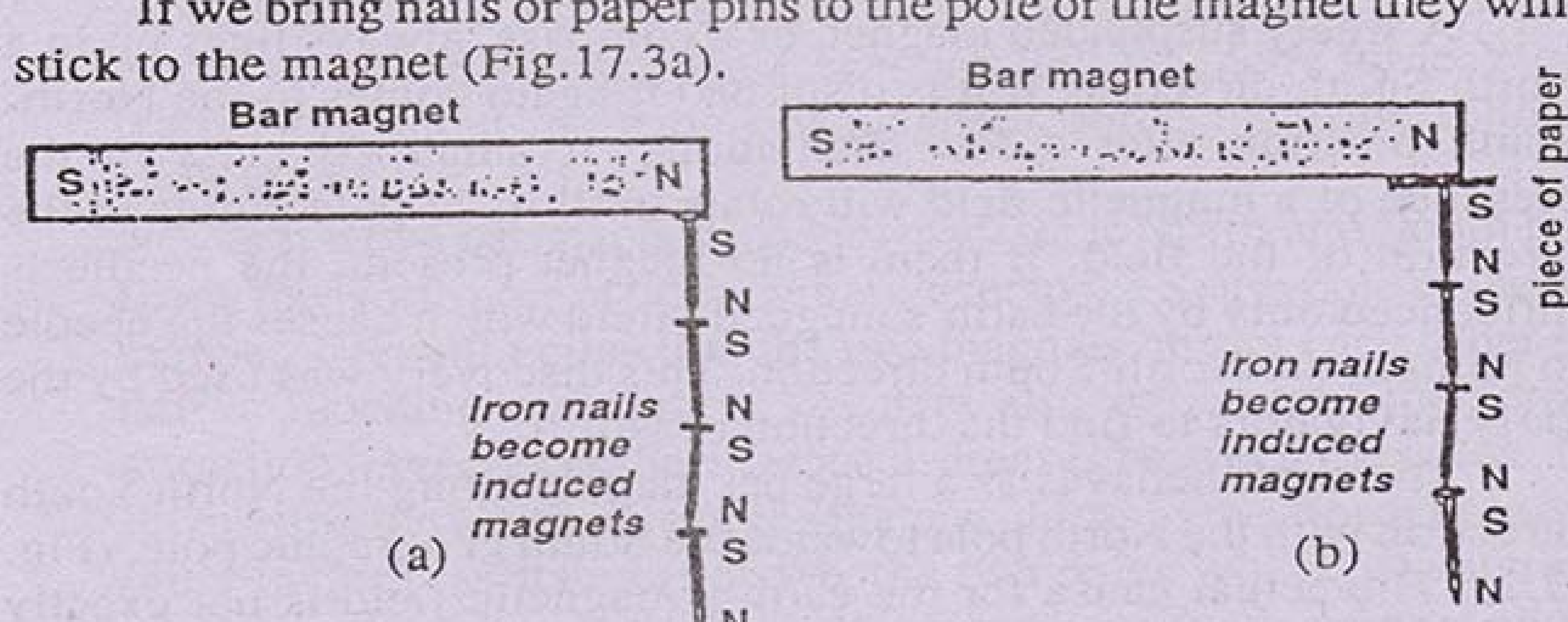


Fig. 17.3 (a) A bar magnet attracts iron nails
(b) A bar magnet attracts iron nails without direct contact

We have noted that the nails were not magnets before they were exposed to the bar magnet and there was no force of attraction or repulsion among themselves. When the nails are removed away from the magnet they again do not attract each other and do not retain their magnetization. Now repeat the experiment again with one change. If we now place a small piece of thin paper between the first nail and the bar magnet. The chain of nails still can be formed without a direct contact with the pole or the bar magnet (Fig. 17.3b).

A substance which behaves like a magnet in the presence of a strong field is called ferromagnetic. The ferromagnetic substances which become magnets in the presence of a magnetic field and lose their magnetism when removed from the magnetic field are called soft ferromagnetic substances e.g, soft iron. Iron may be alloyed with certain materials such as aluminium, silicon and carbon. Once the alloys, such as steel, are magnetized they retain their magnetization. These are called *hard ferromagnetic substances.*

17.7 METHODS OF MAKING MAGNETS

(i) Single Touch Method

If we take a hard steel bar and rub it with one end of a magnet in the direction from S to N, keeping the magnet in an inclined position as shown in Fig. 17.4. On reaching the end N, the magnet is lifted and the same end is brought back to the end S of the bar. The process is repeated several times.

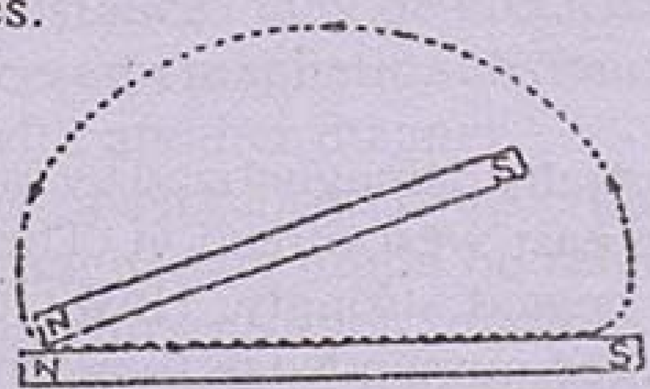


Fig. 17.4 Making a magnet by single touch method

The bar SN will then be magnetized. The end S will have the same polarity as that of the rubbing pole of the magnet and the end N will have polarity opposite to that of the rubbing pole. Each time the bar is stroked at the end S, the angle of inclination should remain the same. Now check your magnet for polarity and strength.

(ii) Electrical method

The best and quickest method of making magnets is the electrical method. Here the magnetic effect of an electric current is utilized. A steel bar in a U-shape is wound with an insulated copper wire as shown in Fig. 17.5. Making sure that the two core arms are wound in opposite directions. The coil is connected to a battery and strong current is passed through the coil for a few seconds.

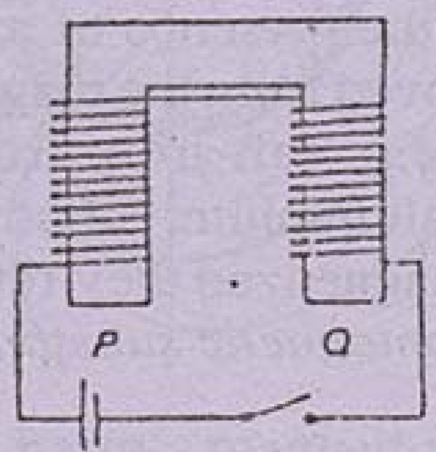


Fig. 17.5 Electrical method of magnetization

In a similar way a bar magnet can be magnetized by putting it inside a solenoid and passing a current through the coil. The specimen will be magnetized. The direction of current in the coil determines the polarity of the magnet. We will discuss this in the section dealing with the field of a current carrying wire.

Demagnetization

Magnets can be partially demagnetized by hammering them when they are pointing in the east west direction. They can also lose their magnetism if they are heated strongly. However, these methods of demagnetization are not recommended as they tend to damage the steel or iron bars. The most efficient method of demagnetizing a magnet is to use an alternating current. The magnet is placed in a solenoid through which an alternating current is flowing. A 12 volt transformer may be used to supply the required current from the main power line. It is recommended that the solenoid be placed in the east-west direction and the magnet is pulled out while the current is still flowing.

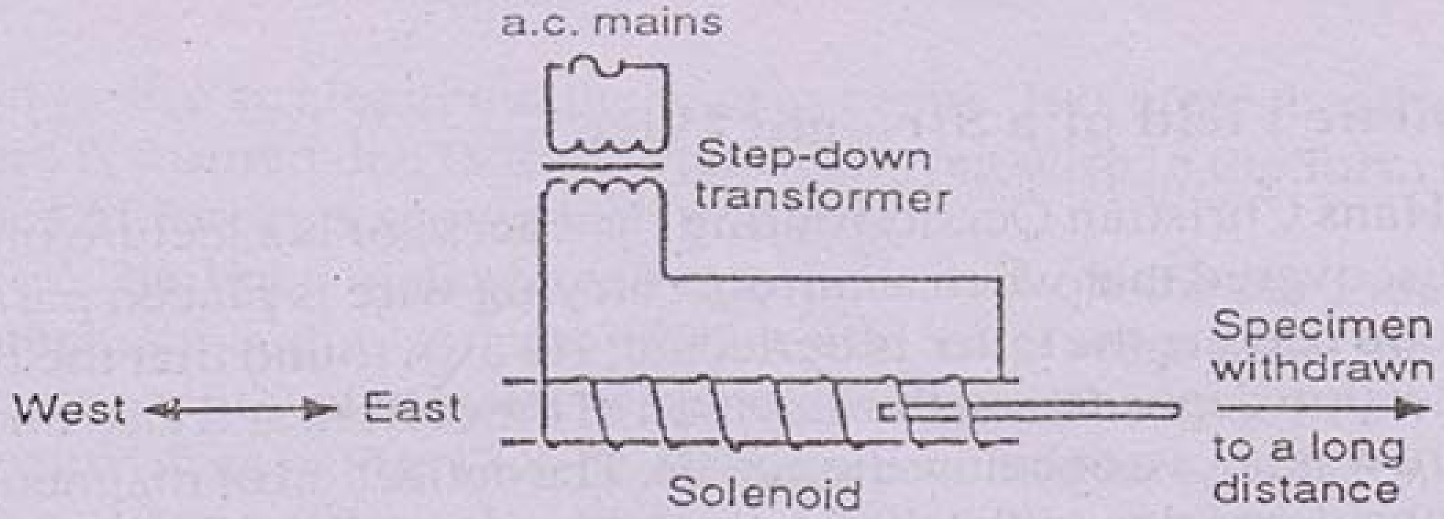


Fig. 17.6 Demagnetization

The alternating current reverses direction at a rate of 100 time per second and hence causes the magnetism of the material to reverse the polarity at the same rate. As a result of this rapid reversal the magnetization is gradually diminished (Fig. 17.6).

17.8 MAGNETIC EFFECT OF CURRENT

The facts about magnetism, that you learned in the previous sections, were compared to the properties of static electricity. It is noticed that there is a remarkable similarity between magnetism and static electricity. These are given below.

(a) There are two types of charges, positive and negative. There are two types of magnetic poles, North and South.
(b) Like charges and like poles repel each other, opposite charges and opposite poles attract each other.
(c) Charged objects set up electric fields of force; the magnetic objects set up magnetic fields of force.
(d) Certain substances may be electrically charged by rubbing together, certain magnetic substance may be magnetized by rubbing with another magnet.

These and other observations caused the early scientists to think that electricity and magnetism might be related phenomena but until the end of the 18th century no such relationship was found.

In 1820 Christian Oersted discovered that an electric current produces a magnetic field in its neighbourhood. This was an important discovery to relate electricity and magnetism and a new branch of physics called *electromagnetism* was born. When an electric current passes through a straight wire the magnetic field created consists of field lines in concentric circles with the wire at their centres.

Magnetic Field of a Straight Wire

Hans Christian Oersted during the course of his lecture on Simple Cell discovered that when a current carrying wire is placed parallel to a magnetic needle, the latter is deflected. He also found that the direction of deflection depended on the direction of the current and also on whether the wire was above or below the needle. The deflection of magnetic needle was considered due to the magnetic field around the wire.

Ampere was the first to note the presence of magnetic force in a wire in which currents are flowing. If two wires in which currents are flowing in the same direction are placed parallel and close to one another then they will attract one another just like the opposite poles of magnets. If the current is flowing in opposite direction through the wires then there will be force of repulsion between them as shown in figure 17.7. The force between them can not be electric in nature due to the following reasons.

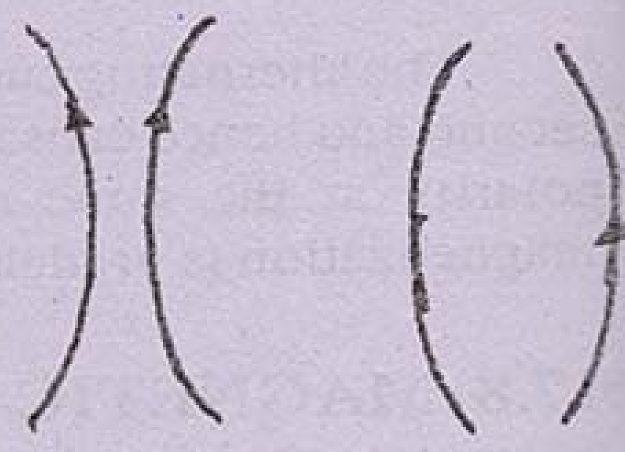


Fig. 17.7

We know that current is produced due to the flow of electric charge. The number of electrons and protons in a conductor are always equal in both the cases when the current is flowing or not flowing. The electric field of positive charge always remains neutralized by the field of negative charge in the wire. Moreover by reversing the direction of current in a wire, the nature of electric charge can not be changed to convert force of attraction into force of repulsion between the two wires. The force disappears when current in one of the wire is stopped. These observations indicate that electric current produces a field around a wire which is not an electric field but it shows magnetic force. "Hence a magnetic field is produced around a wire in which current is flowing"

Since an electric current has magnetic effect in a wire, we should expect it to be surrounded by magnetic lines of force. The following experiments show that circular lines of force are formed around a current carrying wire.

Experiment: A wire is passed vertically through a hole in a cardboard. The two ends of wires are connected to the terminals of a cell. Fine iron filings are sprinkled on the cardboard. The current is switched on and the cardboard is tapped gently. The iron filings set in a series of

concentric circles about the wire as centre. It is clear that the magnetic field is formed due to the current carrying wire in the form of circular lines of force as shown in Fig. 17.8(a).

The lines of force can also be traced on the cardboard using a compass needle. A compass needle is placed near the wire and direction indicated by its north pole is noted. The lines of force are anti-clockwise as in Fig. 17.8(b). If the current is reversed by changing over the cell connections the compass needle will swing round and point in the opposite direction. The direction of lines of force will now be clockwise as shown in Fig. 17.8(c).

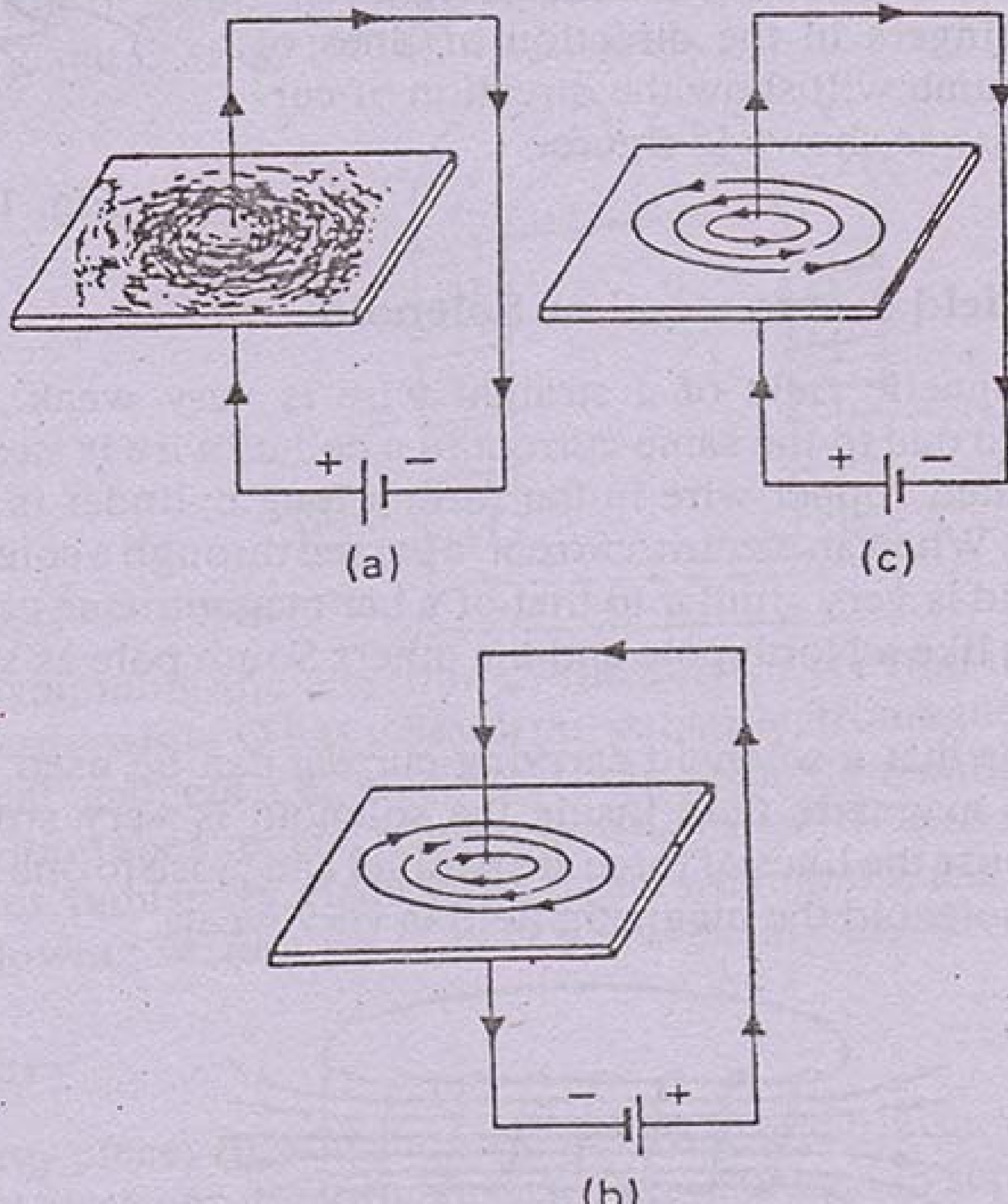


Fig. 17.8

James Clark Maxwell gave a rule relating the direction of the magnetic lines of force round a wire to the direction of the current flowing through it. This is known as the RIGHT HAND RULE and it is stated as follows:

"Imagine the wire to be grasped in the right hand with the thumb pointing along the wire in the direction of the fingers will give the direction of the magnetic lines of force"

Right hand rule is used to know the direction of magnetic lines of force if the direction of current is known, it can also be used to find the direction of current in a wire, if the direction of lines of force is known as shown in Fig. 17.9.

"Hold the wire in the extended right hand curling the fingers in the direction of lines of force. The thumb will show the direction of current in the wire as shown in figure.

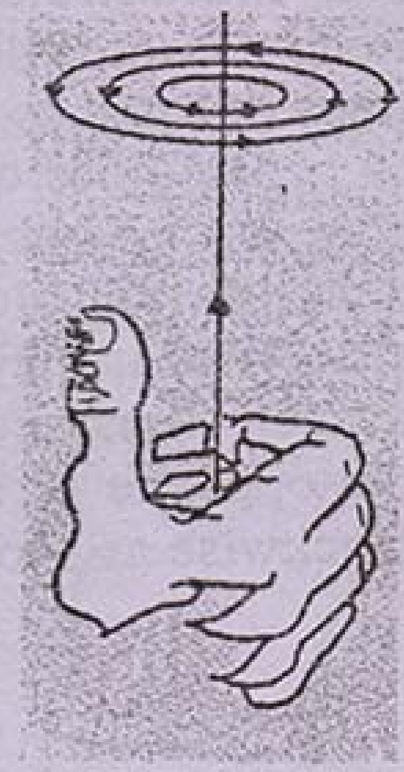


Fig. 17.9

Magnetic Field due to a coil or Solenoid

The magnetic field of a straight wire is very weak. But the magnetic field due to the same current in a coil of wire is stronger. A coil of insulated copper wire in the form a long cylinder is called a SOLENOID. When an electric current is passed through a solenoid the magnetic field is very similar to that of a bar magnet. One end of the solenoid acts like a North pole and the others South pole as shown in Fig. 17.10.

It means that a solenoid carrying current can be used as a bar magnet. The magnetic field inside the solenoid is very strong and uniform because the lines of force are parallel and close to one another. Outside the solenoid the magnetic field is very weak.

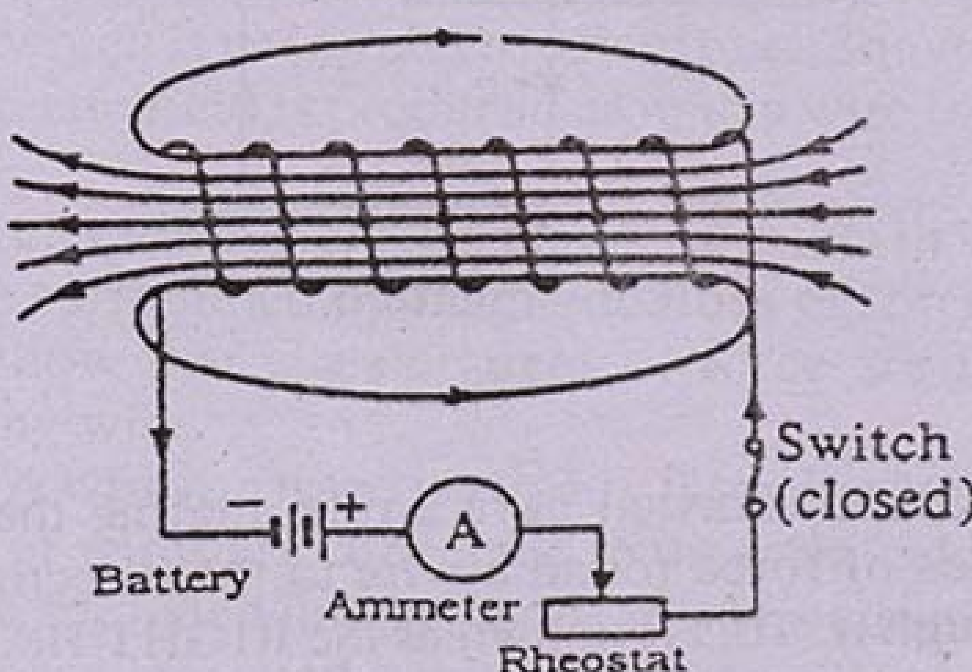


Fig. 17.10

17.9 ELECTROMAGNETS AND THEIR APPLICATIONS

We have already seen that a coil, or a solenoid acts like a bar magnet. If soft iron is inserted in the core of the solenoid the magnetic field due to the current in the solenoid is multiplied by thousands. When the current is switched off there is no magnetic field present. Such a magnet which can be energized by an electric current is called an *electromagnet.* Beside electricity it is the electromagnet which has played an important role in the life of modern man, thanks to Oersted's discovery. One form of the electromagnet is a U-shaped soft iron bar which has coils of insulated wire wound in opposite directions on each of its arms P and Q (Fig. 17.11).

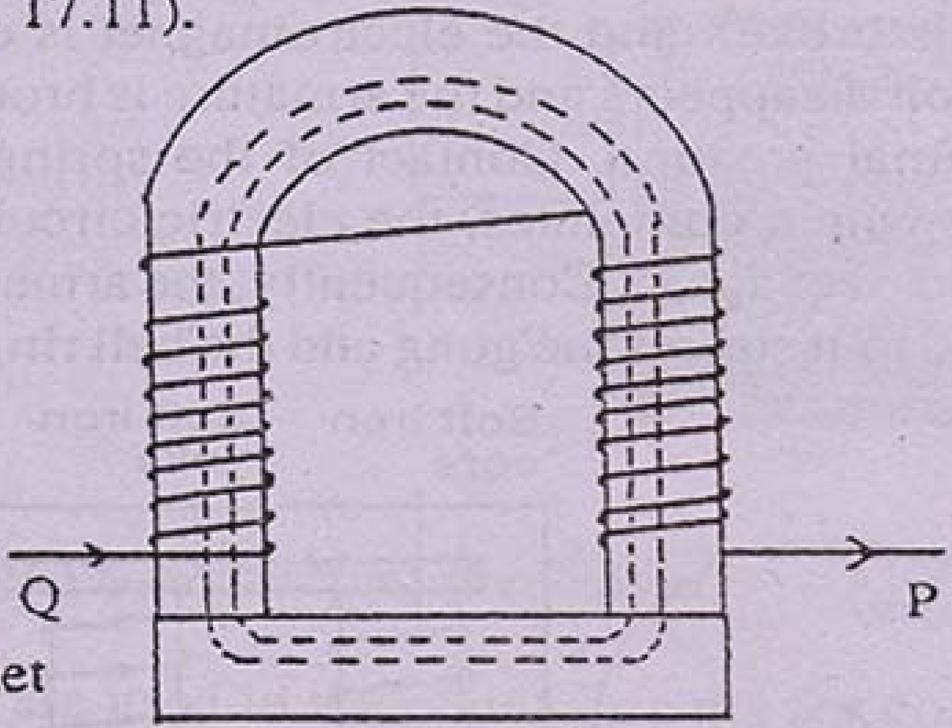


Fig. 17.11 Electromagnet

By examining the direction of the current we find that P has a Norlh polarity while Q has a South polarity. Electromagnets are used in electrical gadgets i.e. electrical bell, induction coil, telephone receivers, loud speaker and television. They are also used in scientific laboratories, industries, electric fan, motors, dynamos, and gènerators. In the following section we discuss a few examples.

In industry

Heavy pieces of iron and steel are lifted and transported quickly and safely by strong electromagnets. They are used in separating iron from mixtures containing magnetic and non magnetic substances. Electromagnets are used to produce strong magnetic fields for high power motors and generators and also in material research laboratories.

Electric Bell

The construction and working of an electric bell is given below:

An electric bell consists of an electromagnet. One end of the windings is connected to a terminal T_1 and the other to a spring which is mounted on a soft-iron stripe called "ARMATURE". A rod is attached to the armature, the free end of the rod carries a small hammer which can strike against a bell. A very light spring is attached to a contact adjusting screw which is joined to second terminal T_2 by a wire. The electric circuil is completed through a battery and push switch button connected to the terminal T_1 and T_2 as shown in Fig. 17.12. When the push button is pressed the electric circuit is completed and the armature is attracted towards the electromagnet as a result, the small spring gets detached from the screw due to which the electric circuit is broken and the electromagnet is demagnetised. Hence the attraction disappears and the armature is brought back by the spring to its original position. Contact of the spring with the screw is now remade which completes, the electric circuit. The action is repeated over and over again. Consequently, the armature vibrates and hammer attached to it strikes the gong and the bell rings and sound is produced.

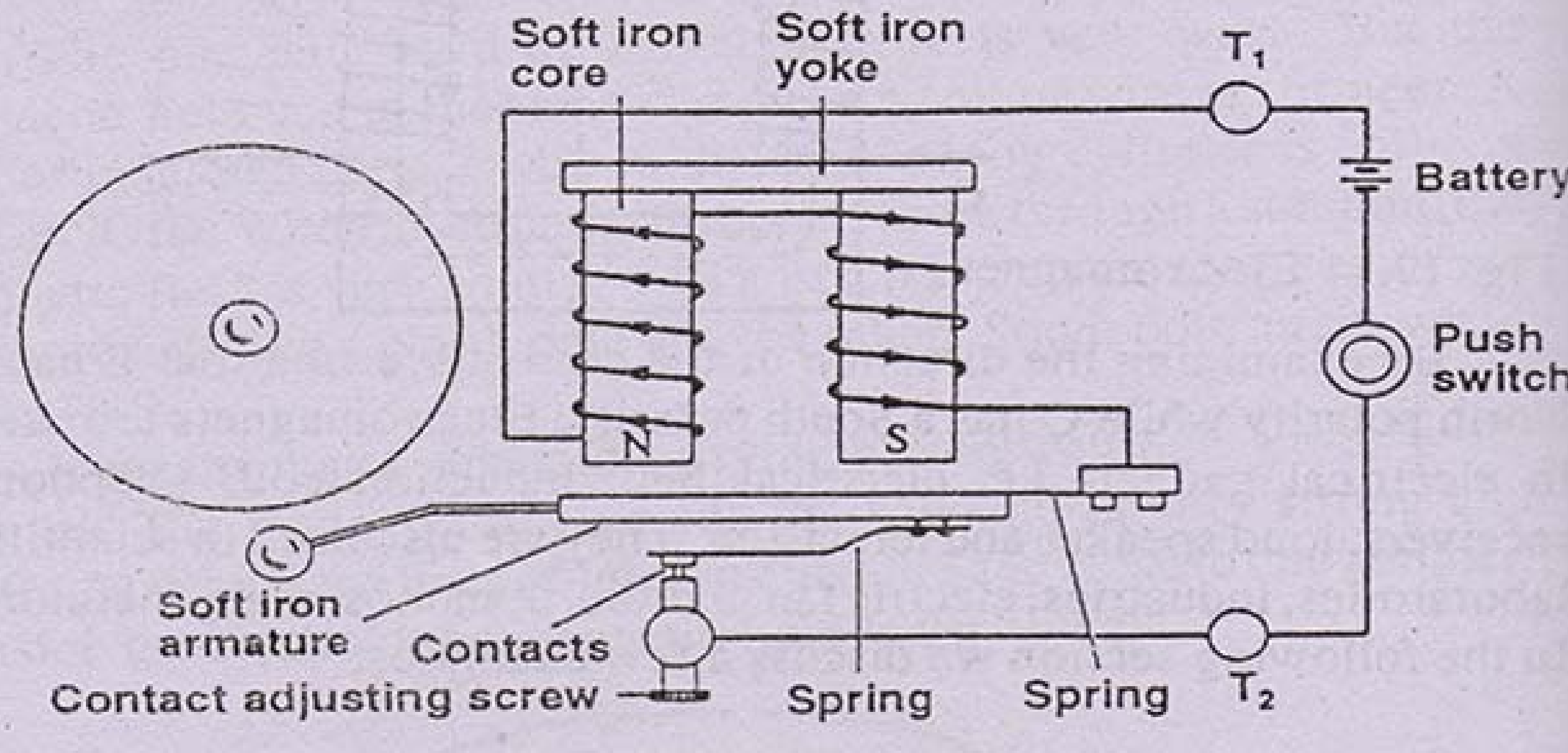


Fig. 17.12 Electric Bell

Telephone Receiver

The telephone receiver is a device which converts electrical energy into sound energy. The ear piece consists of a permanent magnet in contact with two soft iron cores around which coils of insulated wires are wound in opposite directions. A diaphragam made of magnetic alloy is positioned in front of the iron core. When the message is produced from the transmitting end of the line, the sound energy is transformed into electric current and is transported to the ear

piece through the line. The elecIric current so produced varies in magnitude depending upon the frequency of the sound waves. At the receiver, the current passes through the electromagnet coil and energizes the magnet. In this way the magnetic field strength is varied as the current varies. The magnetic diaphragm's pulling force is also varied accordingly. The diaphragm therefore, vibrates and gives rise to sound of the same frequency as that spoken at the other end. (Fig. 17.13).

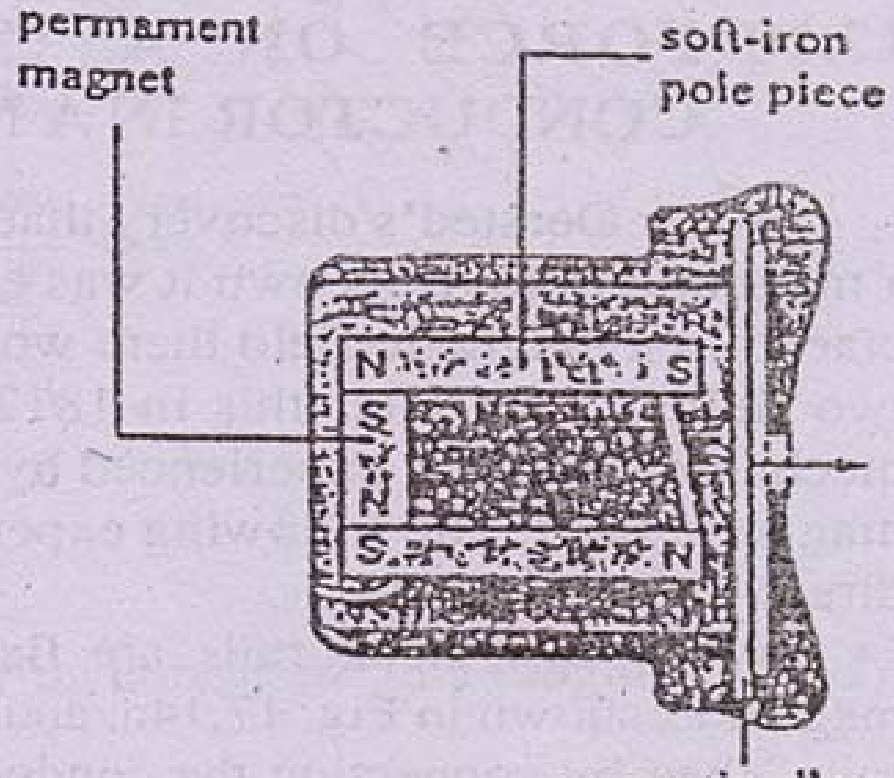


Fig. 17.13 Telephone receiver

Magnetic relays are used in telephone circuits for connecting a line in the telephone exchange to the one from where the number is dialed. Some protective circuits are used in the form of relays when operating with high power electric motors such as a lift operator or in industrial motors and in voltage stabilizers.

Elevated Train

Research is being carried out on the feasibility of using the principle of electromagnets with very fast trains such as the bullet train in Japan. A strong magnetic field can be created to lift the train above the rails so that it appears to float. This removes the frictional force and the train can run very smoothly at high speed. The strength of such an electromagnet can be imagined from the weight of the elevated train. The two electromagnets repel each other with such a strong force that the entire weight of the train is supported and the train is levitated a short distance above the rails. The rails for such trains are made of electromagnets in series on the ground. There are other electromagnets

attached to the lower side of the train. The two magnets, one on the ground and the other beneath the train, are energized such that the like poles face each other. Such tracks are called *Levitation rail roads.* Because of the large current required good conducting wires are needed in these electromagnets. This problem can be solved if super-conducting wires are used instead of normal conductors.

17.10 FORCE ON A CURRENT CARRYING CONDUCTOR IN A MAGNETIC FIELD

After Oersted's discovery that the current in a conductor creates a magnetic field of its own it was expected that if the conductor was placed in a magnetic field there would be an interaction between the two fields. Soon after this in 1812 it was found by Faraday that a mechanical force is experienced by a current carrying conductor in a magnetic field. The following experiment will explain the nature and direction of the force.

When two metal rails are fixed on each side of a horse shoe magnet as shown in Fig. 17.14a, and a conducting rod placed across the rails, then by connecting the conductor to a current source, we notice the rod will move on the rail to the right. If we reverse the direction of the current, the rod also moves in the opposite direction. The rod moves faster if the current is increased. No force is observed and the rod will remain stationary if the magnet and the rod are placed such that the magnetic field is parallel to the length of the rod. If we increase the current there will still be no effect on the rod.

There are several models used by scientists to understand the force on a moving electric charge, or an a conductor carrying a current in a magnetic field. One of these models uses the right hand. The thumb is placed in the direction of the motion of a positive charge or the conventional direction of the electric current. The fingers are placed in the direction of the magnetic field at that position. The direction of the force experienced is then given by the direction in which the palm of the hand would push.

If detailed observation and measurment is carried out it will be found that the force is stronger when the field strength is strong and reduces with a decrease in the field. We conclude the following from the observations.

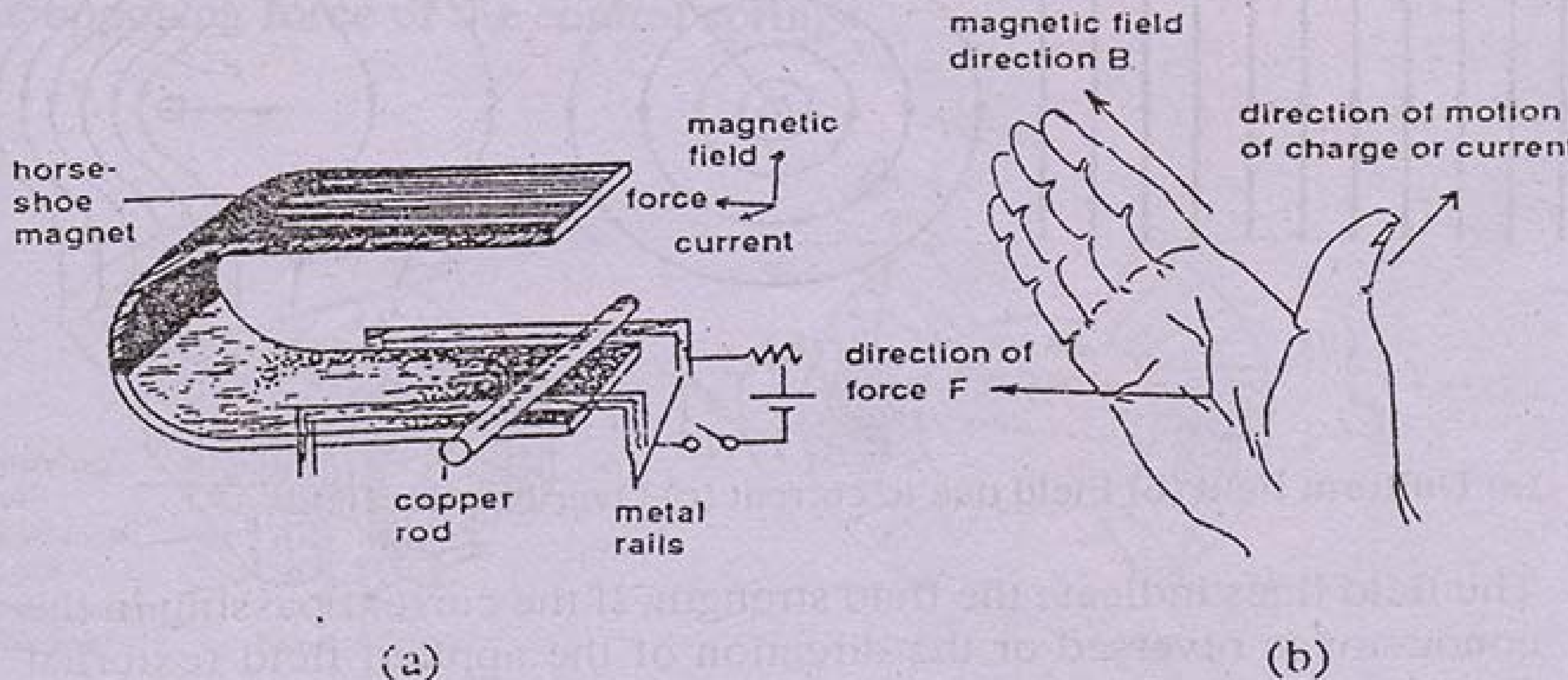


Fig. 17.14 Current carrying conductor in a magnetic field

(i) The current carrying conductor situated in a magnetic field experiences a force whenever it is placed at an angle to the direction of the field.
(ii) The force is always directed perpendicular to the direction of the current and to the magnetic field.
(iii) The magnitude of the force produced, is proportional to the current and the field strength.
(iv) The direction of the force exerted on a current carrying wire in a magnetic field can be determined by using the right hand rule (Fig. 17.14 b).

The direction of the force experienced by the conductor can be explained with the help of the following diagrams.

In Fig. 17.15(a) a uniform magnetic field is shown in which a current carrying conductor is placed. The Fig. 17.15(b) shows the field lines encircling the current in a conductor (flow of charge). In Fig. 17.15(c) the current carrying conductor is shown in the presence of an external magnetic field and the two fields then over-lap. In the region on the left of the conductor the fields are opposing one another and the resultant field strength is weaker. In the region on the right of the conductor the two field reinforce each other thus results in a stronger fields.

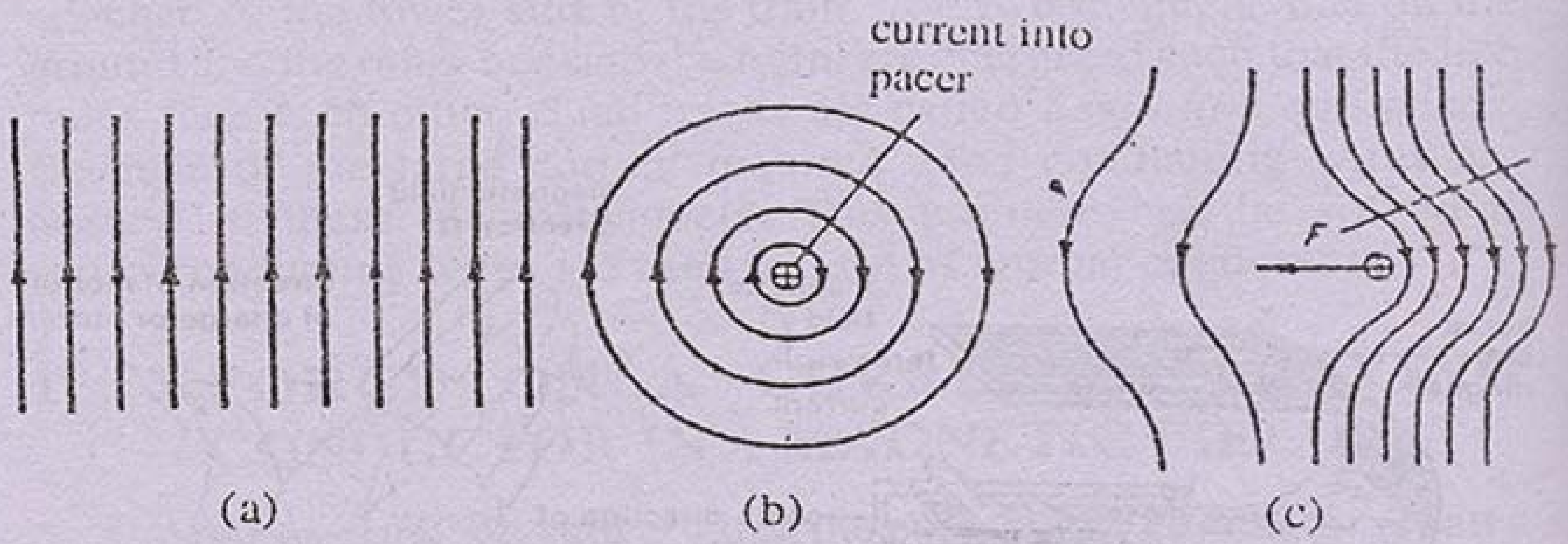


Fig. 17.15
(a) Uniform field (b) Field due to current (c) Over lapping fields

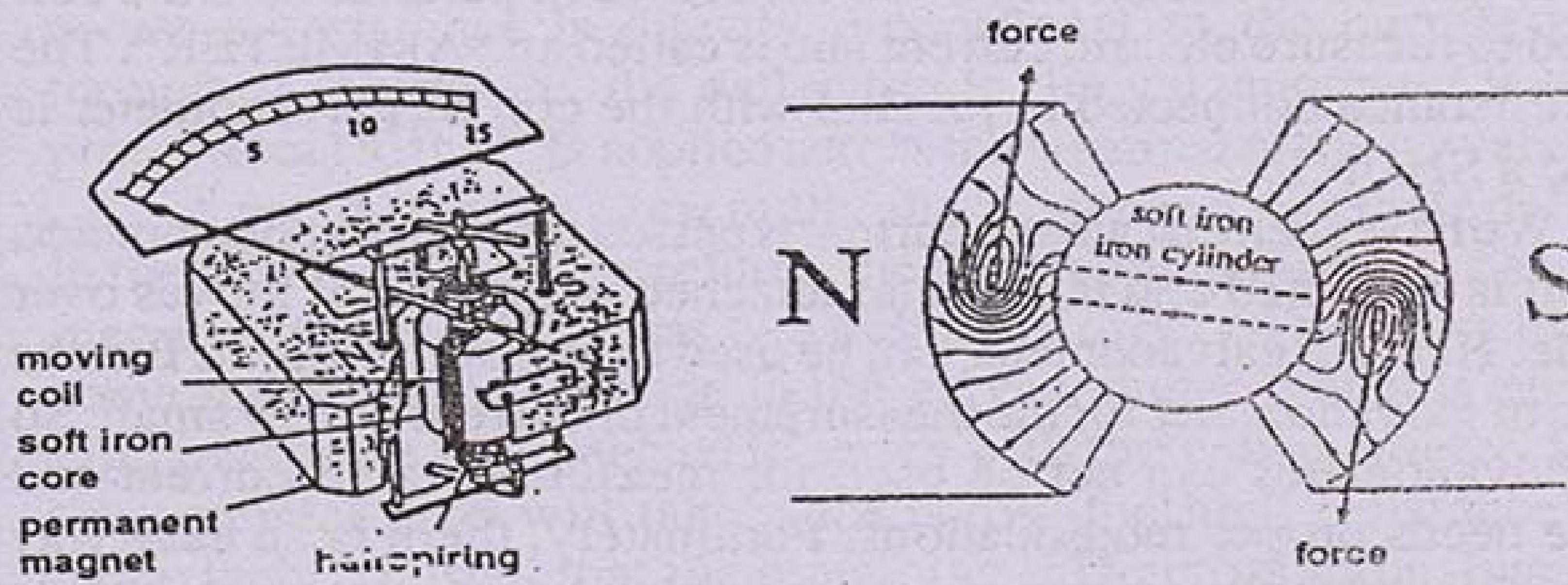
The field lines indicate the field strength. If the current passing in the conductor is reversed or the direction of the applied field (external field) is changed in opposite directions, the force experienced by the conductor will be to the right. The force acts to push the current carrying conductor from the stronger field region towards a weaker field region. The direction of the force is at right angle to both the field lines and the direction of the current.

17.11 MOVING COIL GALVANOMETER

A Galvanometer is a delicate and sensitive device used to measure the magnitude and direction of small currents. Its essential features are shown in Fig. 17.16(a). A rectangular coil of wire is wound on a light frame with a pointer attached on the top. The coil frame is pivoted between the jaws of a large horseshoe magnet. At both ends of the coil hair springs are attached. These springs help in keeping the coil at zero potential and also provide the path for entry and exit to the current. The front and back sides of the coil are parallel to the field of the magnet. The vertical sides of the coil are perpendicular to the field. When current passes through the coil there is no electromagnetic force to act on the front and back wires because they are parallel to the field lines of the magnet. The side wires of the coil will experience forces in opposite directions because the current passes through the side coil perpendicular to the field lines and in opposite directions. On one side the current passes upwards, on the other side it passes down (Fig. 17.16b).

A soft iron cylinder is fixed in the core of the coil to enhance the force on the conductor. The concave shape of the poles combined with

the cylindrical shape of the core creates the radial field to ensure that the field lines are alway perpendicular to the coil. When current passes through the coil the two opposite forces cause the coil to rotate against the opposing force of the control springs.



(a) (b)

Fig. 17.16 The moving coil galvanometer

At one point the electromagnetic force and the controlling force of the spring balance each other and the coil comes to rest. The amount of rotation, measured on the scale, depends on the amount of current flowing through the coil. Also the direction of rotation of the coil depends on the direction of current through the coil. When no current passes through the coil the pointer is adjusted to zero marked at the centre of the scale. When current is allowed to the pointer can rotate to either side depending on the current direction and will measure the amount of current. When the pointer moves to the end of the scale we call it a full scale deflection (F.S.D.) and the amount of current causing the full scale deflection is called the full scale deflection current. The full scale deflection current in most cases is not more than a few milliamperes. That is why we call it a delicate instrument and it must be protected from any current greater than the maximum full scale deflection current.

There are other types of galvanometers where a small magnet is pivoted at the centre of a fixed coil. When current passes through the coil the magnet interacts with the field produced by the current in the coil. The magnet aligns itself with the field and hence rotates from a zero position. When there is no current in the coil the magnet is

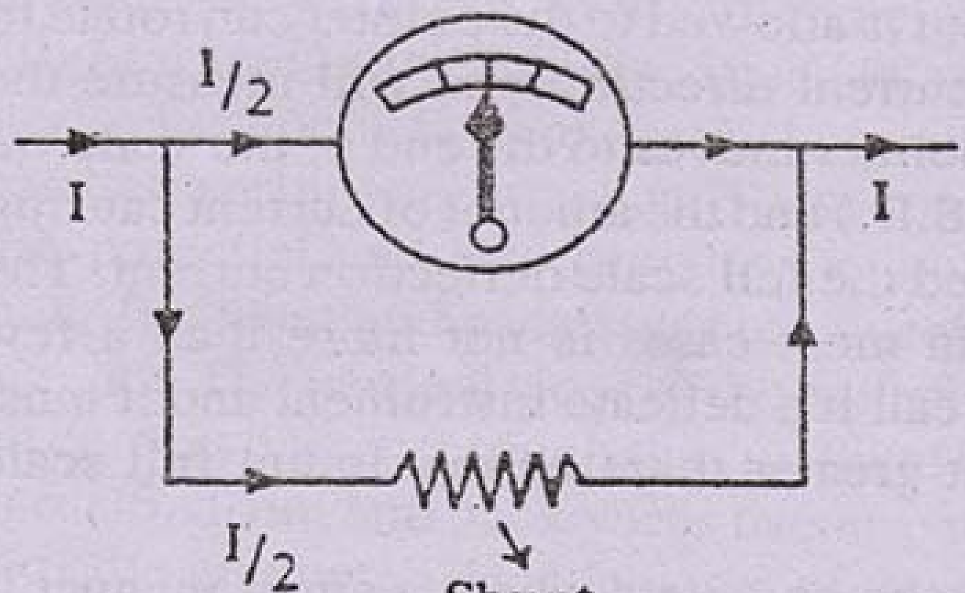
returned to its zero position by a restoring force provided by a spring attached to the magnet. Such galvanometers are moving magnet galvanometers.

17.12 AMMETER

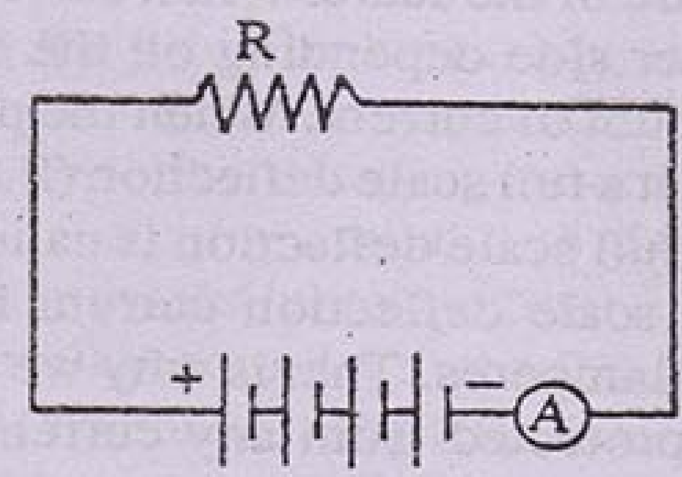
A galvanometer having a low resistance in parallel with the coil is used to measure electric current and is called an "AMMETER". The low resistance connected in parallel with the coil of galvanometer is called a SHUNT.

We have seen that when a current is passed through a galvanometer, its coil is deflected and the pointer attached with the coil moves over a scale. Hence a galvanometer can be used to measure current. But the range of galvanometer for the measurement of current is very small. So the galvanometer can not be used for measuring large current and hence needs proper modifications. Fortunately, there is an easy way out of this difficulty. A low resistance which by passes the great part of the current is placed in parallel with the galvanometer coil. Only a small known fraction of the total current passes through the galvanometer coil. This re resistance acts as a SHUNT, as shown in the Fig. 17.17(a)

An ammeter is always placed in series with other circuit components through which the current is to be measured as shown in Fig. 17.17(b). As ammeter has a low resistance compared with that of the rest of the circuit, therefore it does not introduce unwanted resistance and hence not alter the current passing in the circuit.



(a)



(b)

Fig. 17.17

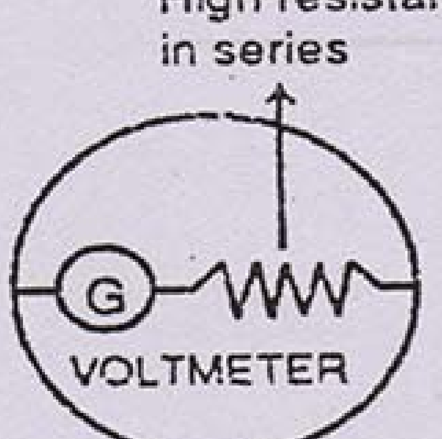
17.13 VOLTMETER

A galvanometer having high resistance in series is used to measure potential difference and is called VOLTMETER as shown in Fig. 17.18(a).

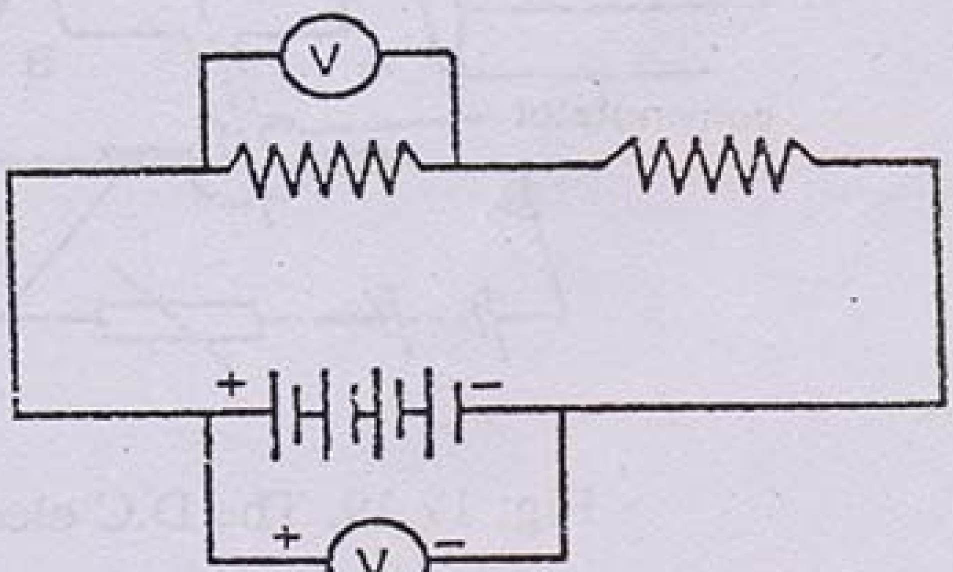
We have seen that the current passing through the galvanometer is proportional to the deflection. By Ohm's law the potential difference across a resistance is directly proportional to the current passing through it. Therefore, the deflection in the galvanometer is directly proportional to the P.D applied across the terminal of the galvanometer.

A small P.D say a few millivolts produces a full scale deflection in the galvanometer. Thereforer, the galvanometer can not be used for measuring high P.D across two points. In order to measure high P.D some modification is made in the galvanometer. For this purpose a high resistance is connected in series with the galvanometer. Most of the potential drop will take place across the high resistance whose value is so adjusted that the remaining P.D across the galvanometer causes current which gives full scale deflection in the galvanometer. The value of resistor connected in series depends upon the range of the voltmeter. For a multi range voltmeter a number of resistors of appropriate values are connected in series with the galvanometer.

When a voltmeter is in parallel with a circuit whose P.D is to be measured, the P.D of the circuit is changed due to the presence of voltmeter. In order to reduce this error, the resistance of the voltmeter should be very high in comparison with the resistance of the circuit. In order to measure the potential difference of a source of e.m.f a voltmeter is always connected in parallel to the source. Similarly to measure the potential difference across a resistor, the voltmeter is connected in parallel to it as shown in Fig. 17.18(b).



(a)



(b)

Fig. 17.18

17.14 SIMPLE ELECTRIC MOTOR

We have already noticed that a current carrying conductor experiences a mechanical force when placed in a magnetic field. This was a key to convert electrical energy into mechanical energy and the devices which converts electric energy into mechanical energy are called electric motors (Fig. 17.19).

Suppose the plane of the coil is initially in a horizontal position. When the current passes from a battery in the direction indicated by the arrows, according to the right hand rule, the arms AD and BC will be acted on by upward and downward forces respectively. The coil will rotate clockwise. When the coil comes to a vertical position, it becomes at right angle to the field and stops. The field produced due to a current in the coil is along the same direction as that of the field of the magnet. According to the right hand rule, the magnetic force on the arm AD acts upward and on the arm BC downward. Hence, there is no rotating force acting on the coil in the vertical position. This has been the case for a galvanometer coil to give a full scale deflection for a 90° rotation. If the coil rotates further in the clockwise direction, a careful examination would reveal that a magnetic force will act to rotate the coil in the anticlockwise direction. However, if we desire lo have a continuous rotation in the clockwise direction, we should change the direction of the field or the current flow on every half rotation of the coil. In most cases the current direction is reversed at each half cycle.

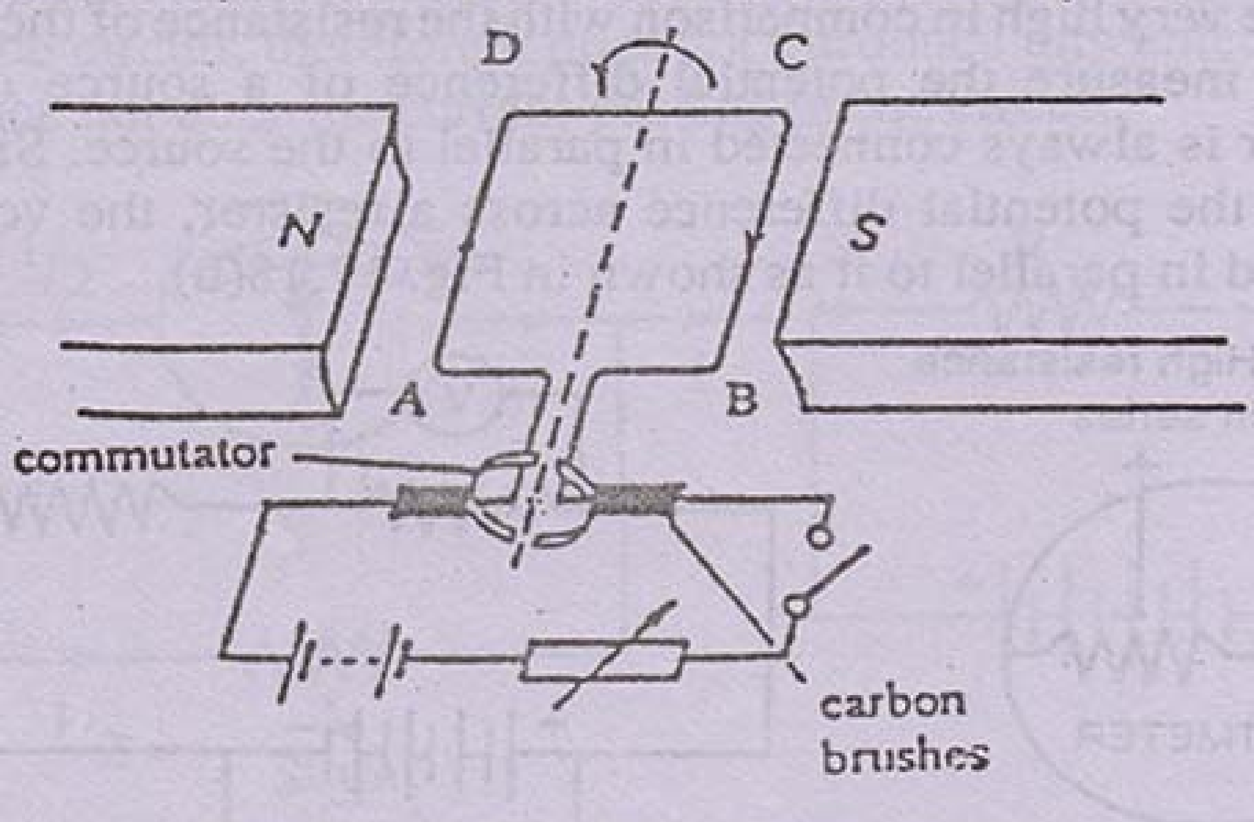


Fig. 17.19. The D.C electric motor

As shown in Fig. 17.19 the ends of the coil are attached to a split copper ring, which rotates along with the coil. This split ring is made

to have a continuous contact with the two stationary pieces of carbon called *brushes.* These brushes are connected to a battery. The current enters through one brush and leaves through the other.

When the coil comes to a vertical position, both the brushes will be at the gaps between the ring segments and no current will flow. However, the coil will continue to move due to its inertia and the ring pieces will come in contact with the brushes again. This time the terminal reverses the direction of the current flow in the coil and causes the arms AD and BC to experience a force downward and upward respectively. Hence the coill continues to rotate in the clockwise direction as long as the current passes through it. The magnetic field arising from the current through the coil changes its direction at each half cycle and thusal ways points in opposite direction to the field lines of the permanent magnet.

Think of a current source which reverses its polarity at each half cycle. Do we need a commutator in a motor for alternating current (A.C)?. The answer is certainl y no and that is the reason a commutator is some time called a *.D.C commutator,* and the motors are called *D.C motors.* The portion which rotates in a motor is called a *rotor or Armature.* The spead of rotatiom of a magnet depends on the following factors

(i) The magnitude of current through the rotor.
(ii) The magnetic field strength of the permanent magnet.
(iii) Number of turns in the coil of rotor.
(iv) Permeability of its armature.

SUMMARY

– Magnets possess two poles, a North pole and a South pole.
– The Earth possesses a magnetic field and acts as a huge bar magnet.
– A substance which behaves like a magnet in the presence of a magnetic field is called a ferromagnet.
– Magnets may be made by (a) single touch method (b) electrical methods. They may be demagnetized by heat, rough treatment or the application of an alternating magnetic field.
– Current produce magnetic field. The direction of the field is given by the right hand rule.
– If a current carrying wire is been into the shape of a coil it will act as a bar magnet.
– Electromagnets have many applications e.g., bells, telephones, relays and elevated trains etc.
– A current carrying conductor in a magnetic field experiences a force.
– A moving coil galvanometer is a sensitive device to measure the magnitude of very small currents.
– A galvanometer can be converted to an ammeter by the addition of a small shunt resistance in parallel with it.
– A galvanometer can be converted to a voltmeter by the addition of a high resistance in series with it.
– The mechanical force on a coil can be translated to rotar motion in a device called an electric motor.

QUESTIONS

17.1 Write answers to the questions given below:

(i) Explain what is meant by magnetic field.
(ii) Explain the right hand rule for the magnetic force.
(iii) Explain the working of an electric bell.
(iv) Give the principle, construction and working of a moving coil galvanometer.
(v) How can a galvanometer be converted into voltmeter and ammeter.
(vi) Explain the working of an electric motor.

17.2 Fill in the blanks.

(i) A freely suspended magnet always points in the __________ direction.

(ii) The region around a magnet where its magnetic effect is felt is called its __________.

(iii) The magnetic lines of force __________ longitudinally.

(iv) The magnetic line of force never __________ one another.

(v) If an electric current is passing through a wire then a __________ is produced around it.

(vi) If the electric current is flowing from top to bottom in a wire, then the direction of the lines of force will be __________.

(vii) If the flow of current at any end of a coil is anti clockwise then this end will be a __________ pole.

(viii) __________ is a sensitives electrical instrument that detects the presence of current in a circuit.

(ix) Ammeter is an electrical instrument used for measuring __________.

(x) The instrument used to measurs the potential difference is known as __________.

17.3 Given below are a few possible answers to each statement, of which one is correct. Identify the correct one.

(i) Like poles of a magnet __________ each other.
(a) attract (b) repel (c) neither attract nor repel (d) sometimes attract and sometimes repel.

(ii) The relation between electric current and the magnetic field was discovered by __________.
(a) Newton (b) Faraday (c) Fleming (d) Oersted.

(iii) If a current is flowing through a solenoid, then the north pole of the solenoid can be found using __________ rule.
(a) right hand (b) left hand (c) Faraday's (d) Lenz's

(iv) A galvanometer can be converted into an ammeter by connecting a wire of low resistance __________ with the galvanometer.
(a) in series (b) in parallel (c) in a combined way (d) in no way

(v) To measure current in a circuit an ammeter is always connected __________.
(a) in series (b) in parallel
(c) in any way (d) parallel to voltmeter.

17.4 Pick out true and false from the following sentences.

(i) The magnetic lines of force start from the south pole and end at north pole and they are continuous through the body of the magnet.

(ii) The magnetic lines of force can pass more easily through iron than air.

(iii) Faraday discovered a connection between electric current and magnetism.

(iv)) The magnetism of an electromagnet depends upon the number of turns and the magnitude of the current passing through the coil.

(v) The magnetic field produccd by a current passing through a solenoid is similar to a magnetic field due to a bar magnet.

(vi) If a current carrying conductor is held perpendicular to a uniform magnetic field, then no force acts upon it.

(vii) The working of a galvanometer depends upon the magnetic field of a permanent magnet and its interaction with the magnetic field produced around the current carrying wire.

(viii) A galvanometer having a shunt fitted parallel to it is called a voltmeter.

(ix) A galvanometer can be converted into an ammeter by connecting a wire of suitable resistance in series with it.

(x) A good voltmeter is that voltmeter whose resistance is so high that a very small current passes through it.

CHAPTER - 18

ELECTRONICS

18.1 INTRODUCTION

Electronics is a branch of physics which deals with the development of electron emitting devices and, their utilization and controlling of electron flow in electrical circuits designed for various purposes. The impact of electronics on the daily life of people all over the world is considerable now a days. Radio, television, stereo hifi sound systems, motion pictures and video cassette recorders provide a lot of entertainment and information. Similarly automatic washing machines, microwave ovens, robots, telephone systems and pocket calculators have made jobs simple and convenient. Electronic computers are being used in business, offices, industry, hospitals and research centres. Electronics also controls the operations of satellites orbiting around the earth. These satellites are designed to serve as worldwide communications, networks, to scan earth's natural resources and to collect data on weather and climate.

18.2 n-type AND p-type SUBSTANCES

Highly pure silicon and germanium crystals are almost complete insulators, especially at low temperatures. This is so because all four valence electrons of each atom form covalent bonds with their neighbouring atoms, as shown in Fig.18.1. Since these electrons are tightly held in covalent bonds there are no really free electrons to form an electric current on the application of a potential difference. The electrical conductivity of silicon or germanium can however, be increased by adding in them a very small amount of an element which has either three or five valence electrons in its atoms. This process is called doping.

If a silicon crystal is doped with a pentavalent element such as arsenic (Fig.18. 2a), four out of five valence electrons of an atom form four covalent bonds with four neighbouring silicon atoms. The fifth valence electron is however, free to move about, which makes the doped silicon crystal a better conductor. We call this material an n-type

semiconductor because there is an excess of negative electrons in it to form an electric current.

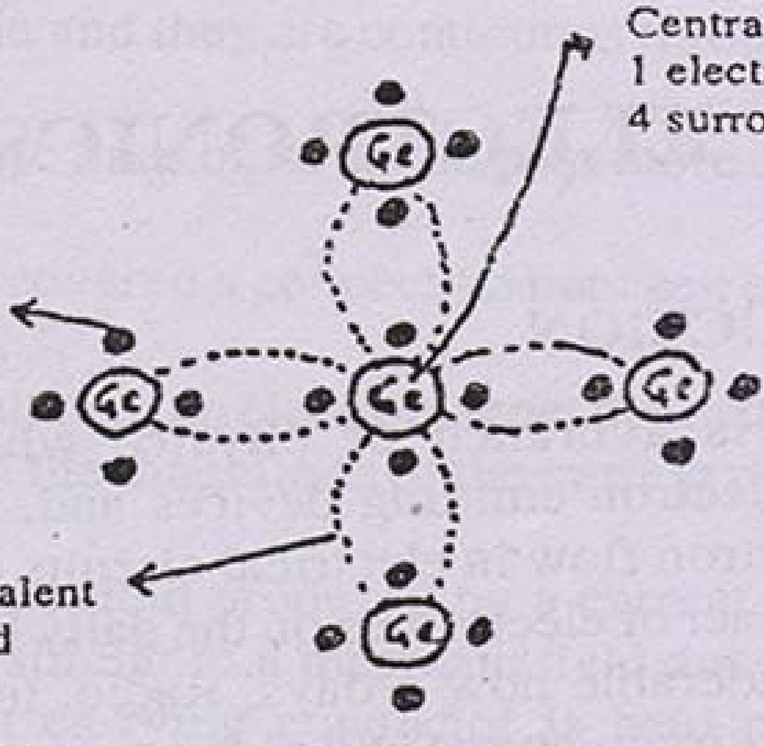


Fig. 18.1 Bonding of germinium atoms

If a silicon crystal is, however, doped with a trivalent element such as indium (Fig. 18.2b), all the three available valence electrons of an atom form three covalent bonds with

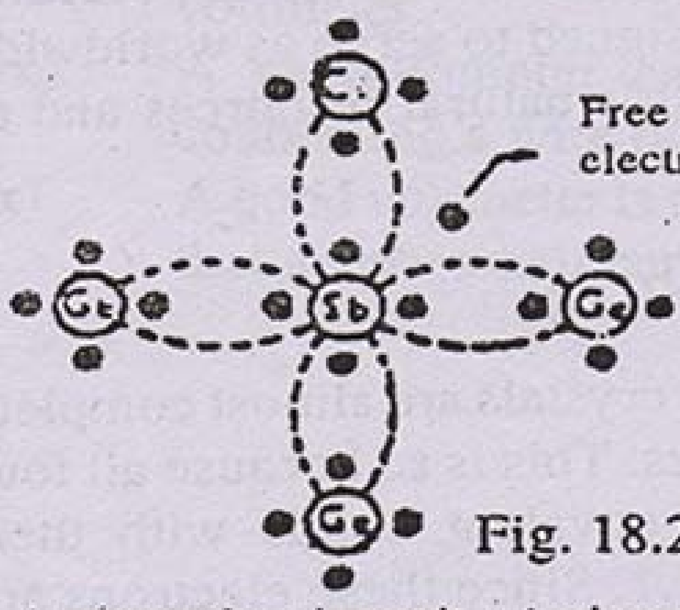


Fig. 18.2 (a)

A pentavalent impurity. Anti mony added to the germanium crystal furnishes free electron, making n-type material

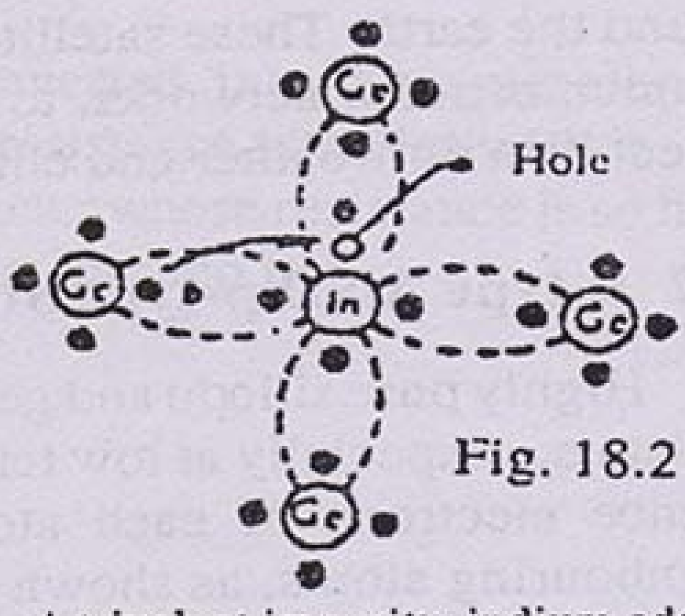


Fig. 18.2 (b)

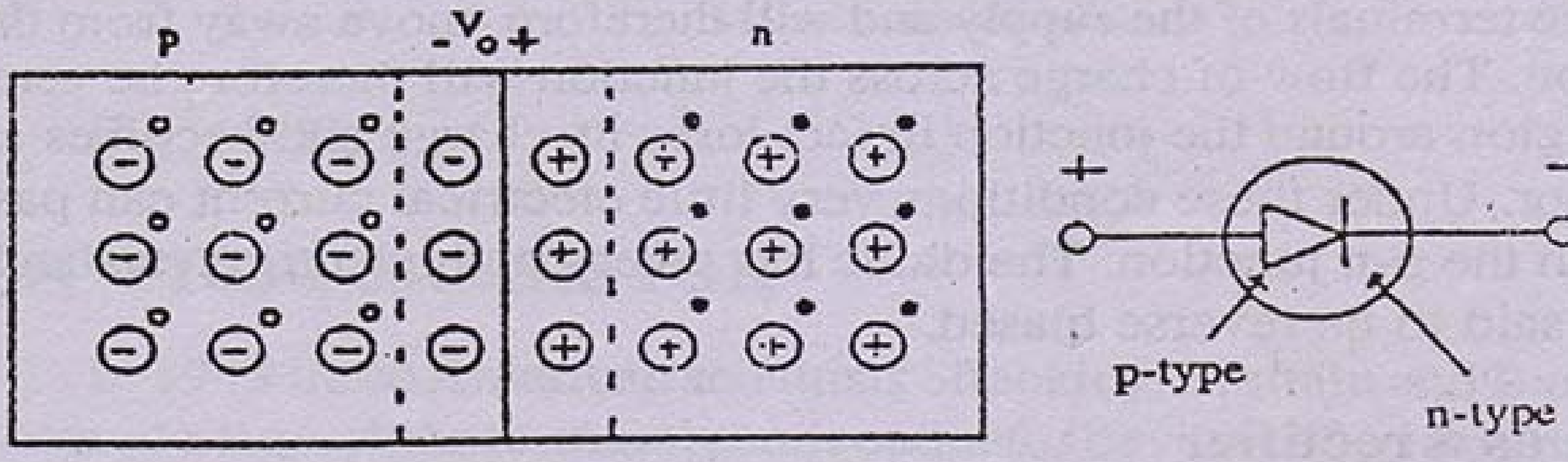
A trivalent impurity. indium added to the germanium crystal furnishes hole. making p-type material.

three neighbouring silicon atoms. A space, called a hole, is therefore left in the silicon crystal due to the shortage of an electron. This hole behaves like a positive charge, and can move from place to place in the crystal on the appli ation of potential difference. Such a material is called a *p-type semiconductor* because there is an excess of positive charges in it. It is worthy to note that holes flow in the direction (opposit) to electrons. Germanium can also be doped with pentavalent or trivalent elements (as an impurity) to form n-type or p-type material respectively.

18.3 p-n JUNCTION DIODE

A p-n junction diode is an electronic device formed from a p-type and an n-type semiconductors. A p-n junction is fabricated by placing a small amount say of indium, on a plate (called wafer) of n-type germanium. Indium on heating to about 550°C melts and diffuses, through a small part of the n-type germanium. Indium being the p-type impurity converts the part of the n-type germanium to p-type material. Thus, a junction forms between a p-type section and an n-type section of the germanium. A brass-base is used to fix the p-n junction to which lead are attached for external electrical connections. This is then scaled in a metal or glass container. Since the p-type material has an-excess negative charges (electrons), the electrons from the n-type material and holes from the p-type material flow across the junction and combine. A positive charge layer is formed on the n-type material and a negative charge layer is formed on the p-type material due to the migration of the electrons from n-type to p-type material near the junction. A potential barrier is developed across the junction which prevents further flow of charge from one side to the other.

This is illustrated in Fig. 18.3 a. Such a semiconductor device is called a diode because n and p sides behave as two electrodes and is symbolically represented as shown in fig. 18.3b.



(b) (a)

Fig. 18.3 (a) p-n junction diode (b) diode symbol

Forward bias

When a semiconductor diode is connected to a direct current (D.C) power supply (Battery) as shown in Fig. 18.4a, the p-type material will be at a positive potential and the n-material will be at a negative potential. The positive holes will now drift from the p-type to the n-

type material across the junction while free negative electrons will move from the n-type to the p-type material. the movement of these charges occurs because the positive terminal of the power supply repels holes in p-type material towards the junction, while the negative terminal also repels the free electrons in n-type material towards the junction. Thus an electric current can easily pass through the p-n junction. The diode has a very low electrical resistance and is said in a condition of forward bias.

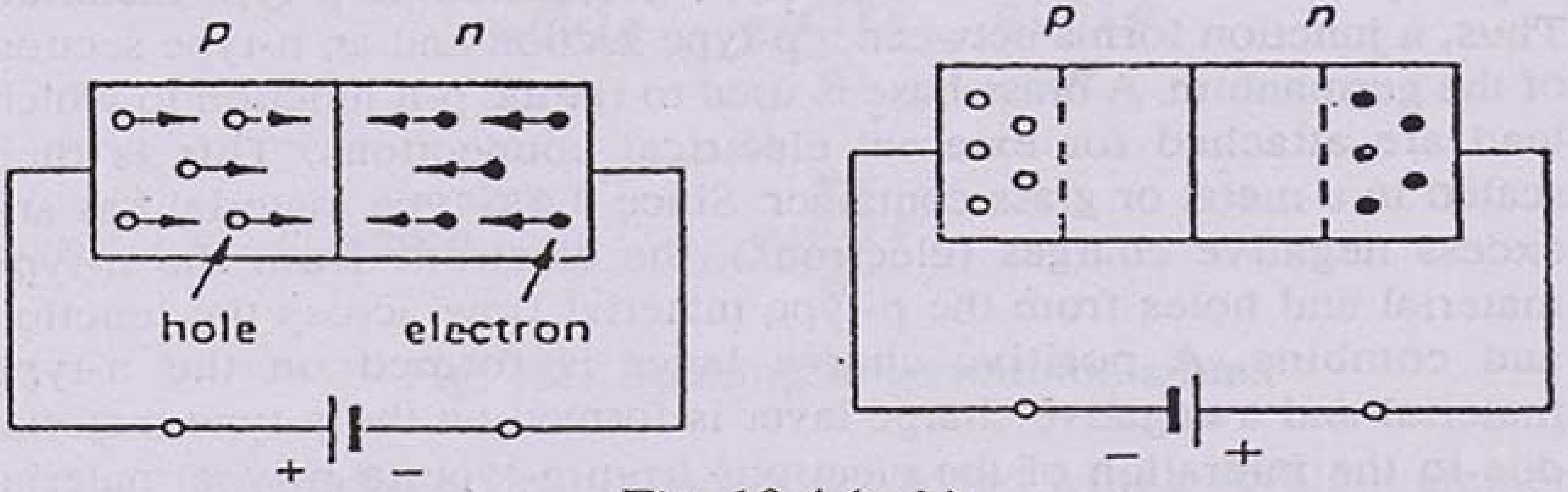


Fig. 18.4 (a, b)

Reverse Bias

If the semiconductor diode is now connected to the D.C power supply as shown in Fig. 18.4b, the holes in the p-type material and the electrons in the n-type material will be attracted by the negative and positive terminals of the supply and will therefore, move away from the junction. The flow of charge across the junction will therefore be zero. The region around the junction in fact loses its charge and becomes an insulator. Under these conditions very little electrical current can pass through the p-n junction. The diode has a very high electric resistance and is said to be reverse biased.

Diode as a rectifier

A p-n junction diode is used as a rectifier because it allows charges to flow in one direction. A rectifier is a device that converts A.C current to D.C current; the phenomenon itself is known as rectification. A circuit that accomplishes rectification is shown in the Fig. (18.5a). The input-voltage V_i is an alternating voltage which has to be rectified. During the positive half cycle of V_i, the p-section of the diode is positive and repells the holes towards the junction and hence resistance of the junction is reduced which causes the current to flow through an external resistance R and an output voltage appears a cross R, On the other hand during the negative half

cycle, the p-section of the diode is negative and it attracts holes which results in the increase of junction resistance. Due to this increase in resistance it does not conduct and no current passes through R, the voltage drop across R becomes zero i. e. $V_o = O$. Hence we see that through a diode only positive half of the A.C cycle passes and rectification is achieved. This is called half-wave rectification. For full-wave rectification we use two diodes in the circuit as shown in Fig.(18.5b).

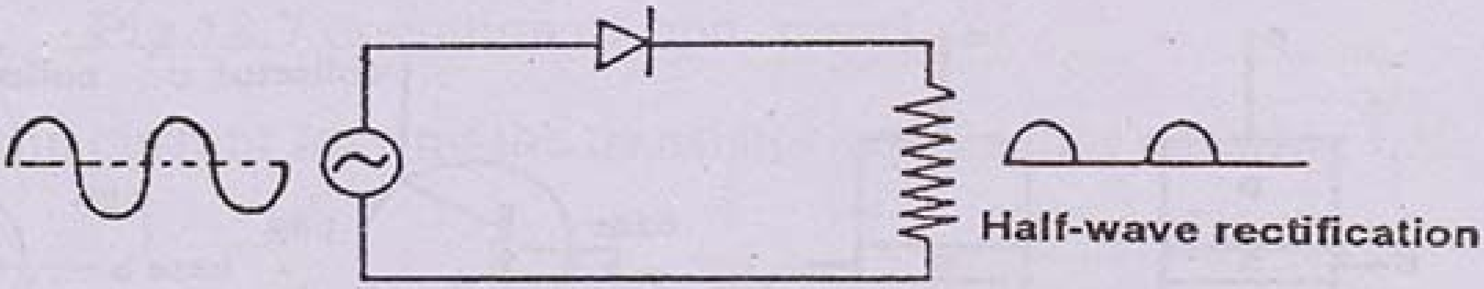


Fig. 18.5 (a)

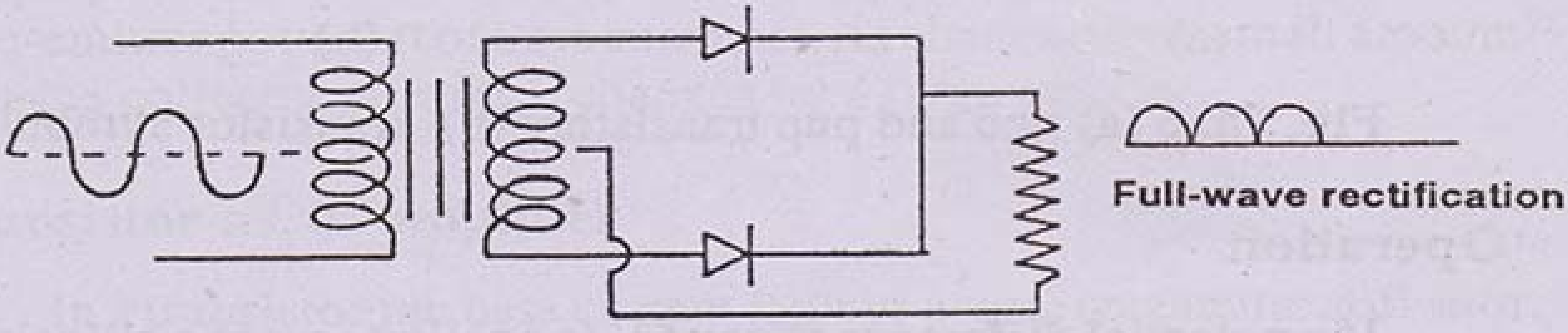


Fig. 18.5 (b)

18.4 THE TRANSISTOR

In 1948 three American scientists Shockley, Brattain and Bardeen invented a tiny, three terminal semiconductor device called a transistor which produced a revolution in the world of electronic devices. Transistors are now being used in radio and television sets audio and video cassette recorders voltage stabilizers, telephone sets, computers and numerous other devices.

Construction

A transistor is basically a semiconductor which consists of a thin (3-5 μm thick) central layer of one type of semiconductor material sandwhiched between two relatively thick pieces of the other type.

There are two types of transistors: npn and pnp. (Fig.18.6 a). The npn transistor has a thin piece of p-type material sandwiched between two pieces of n-type, while the pnp transistor consists of a piece of n-type in between two pieces of p-type. The central part is known as the base (b) and the pieces at either side are called the emitter (e) and the collector (c). In circuit diagrams, the npn and pnp transistors are represented by the symbols given in (Fig. 18.6 b). The arrow on the symbol gives the direction in which conventional (positive) current would flow.

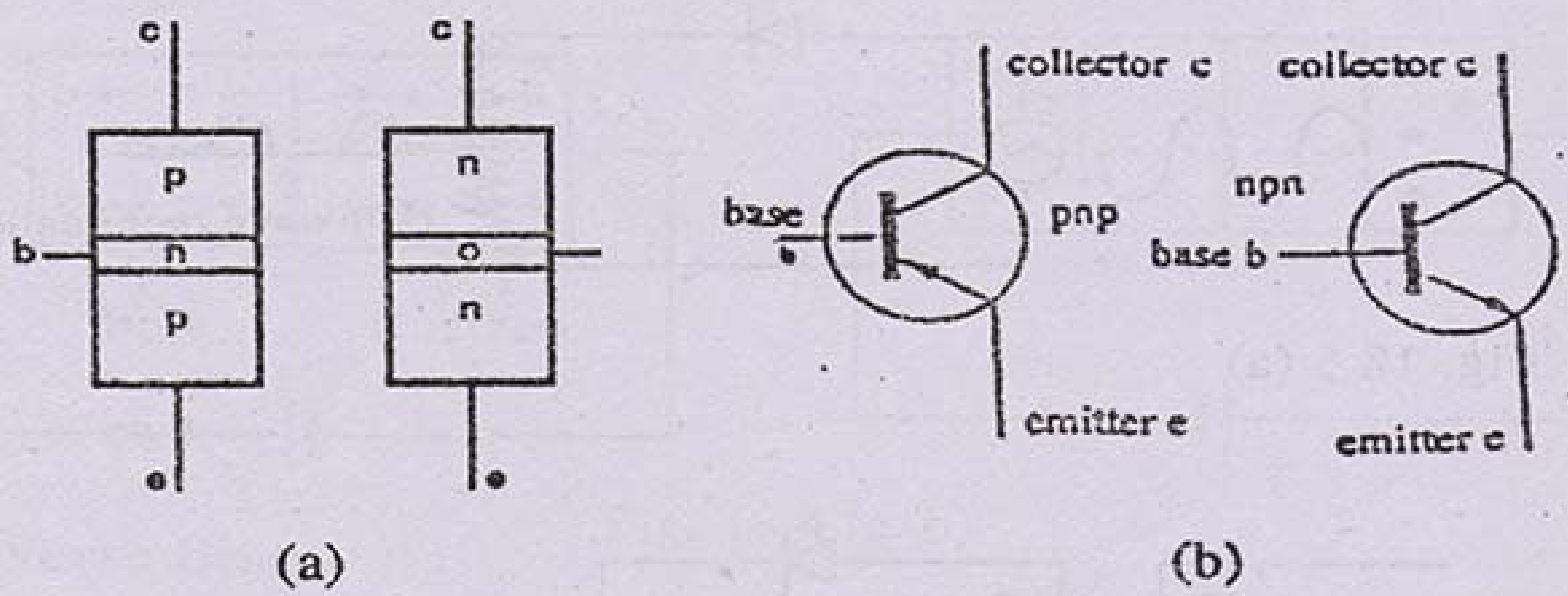
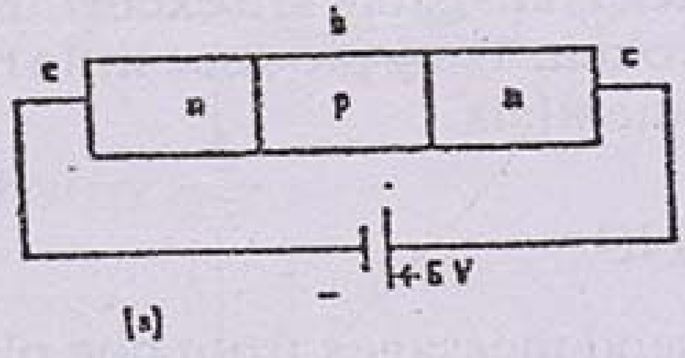


Fig. 18. 6 (a) npn and pnp transistors (b) transistor symbols

Operation

If a potential difference of say 6V is applied across a silicon npn transistor as shown in (Fig. 18.7a) the collector becomes positive with respect to the emitter (the base is left unconnected). The base-collector p-n junction is therefore reverse biased and current cannot pass through the transistor.

Now, if the base-emitter pn junction is forward biased by applying a potential difference of about 0.6 V as shown in (Fig.18.7-b), electrons flow from the n-type emitter into the p-type base. The loss of electrons in the emitter is compensated by electrons entering the emitter from the external circuit and form the emitter current I_e, since the base is very thin and is lightly doped a small number of the electrons from the emitter combine with holes while most electrons pass through the base under the attraction of the positive collector. They cross the base-collector junction and become the collector current I_c in the external circuit. The loss of holes which occurs in the base is compensated by some holes flowing to it from the base power supply. This forms a small base current I_b.



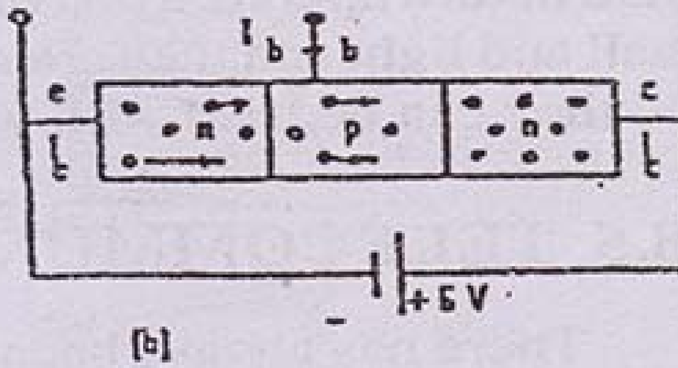


Fig 18.7 operation of npn transistor

Since the current leaving the transistor equals that entering i.e.

$$I_E = I_B + I_C$$

Thus there are two current paths through a transistor. One is the base-emitter path (input) and the other is the collector-emitter path (output). The importance of a transistor is due to the fact that if the base-emitter potential of base current I_b, is changed by a small amount then the collector current I_c changes by a large amount.

Transistor as an amplifier

In a transistor the base current I_b plays a vital role in the collector current I_c. A small change in the base current produce a large change in the collector current. Due to this characteristic a transistor is used as an amplifier in various electronic circuits. A block diagram of the amplifier is shown in (Fig. 18.8).

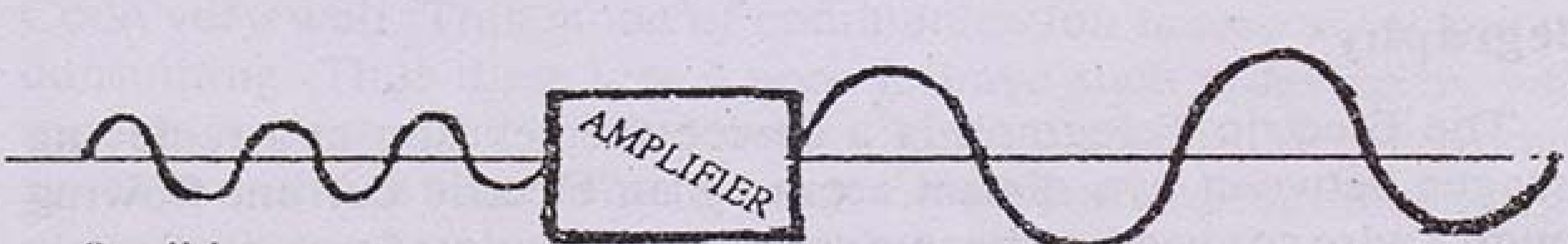


Fig. 18.8 An amplifier changes a small voltage into a large voltage out put.

The transistor is used as a switching device. Anything which can 'switch on' a small base current to a transistor will release a large

collector current to operate a lamp, loudspeaker or relay. The relay could in turn operate a bell or a motor, etc. Transistors are exceedingly small and light and they are not easily broken. They produce little heat and they can be used with very small potentials.

18.5 TELECOMMUNICATIONS

There has always been a need to send messages from one place to another far away by the quickest possible method. In the past, drums were beaten to inform of impending dangers, flags were waved to send messages and at night flashes of light were utilized for this purpose. In the nineteenth century we begin to truely understand electromagnetic theory. According to this theory light and what we now know as radio waves are the same in nature i.e. electric and magnetic field mutually perpendicular to each other travelling at a speed of 3×10^8 m/sec through space, only the frequencies are different. Man learn to utilize radio waves for sending messages across the globe. The science which deals with the production and proper usage of different frequencies of radio waves for sending sound and visual information is known as telecommunication. In the following sections we shall describe some telecommunication devices.

The discovery of electromagnetic waves in the second half of the nineteenth century helped in a new era of relaying messages from one place to another using these waves. Electromagnetic waves travel at a speed of 3×10^8 metres per second. This is called telecommunication. In the following sections we shall describe same telecommunication devices.

Telegraphy

The Electric Telegraph is a device for sending and recieving messages between two distant areas by an electric current flowing through a wire connecting the two areas in its simplest form, an electric telegraph consists of an electric battery connected through a tapping key (called the sender) to an electric buzzer (called the receiver) as shown in Fig. 18.9. Only one wire is needed between the sender and the receiver as the circuit is completed by connecting their other ends to the earth, which being moist a few feet below the surface, acts as a good conductor. when the tapping key is pressed the receiver produces a buzzing sound.

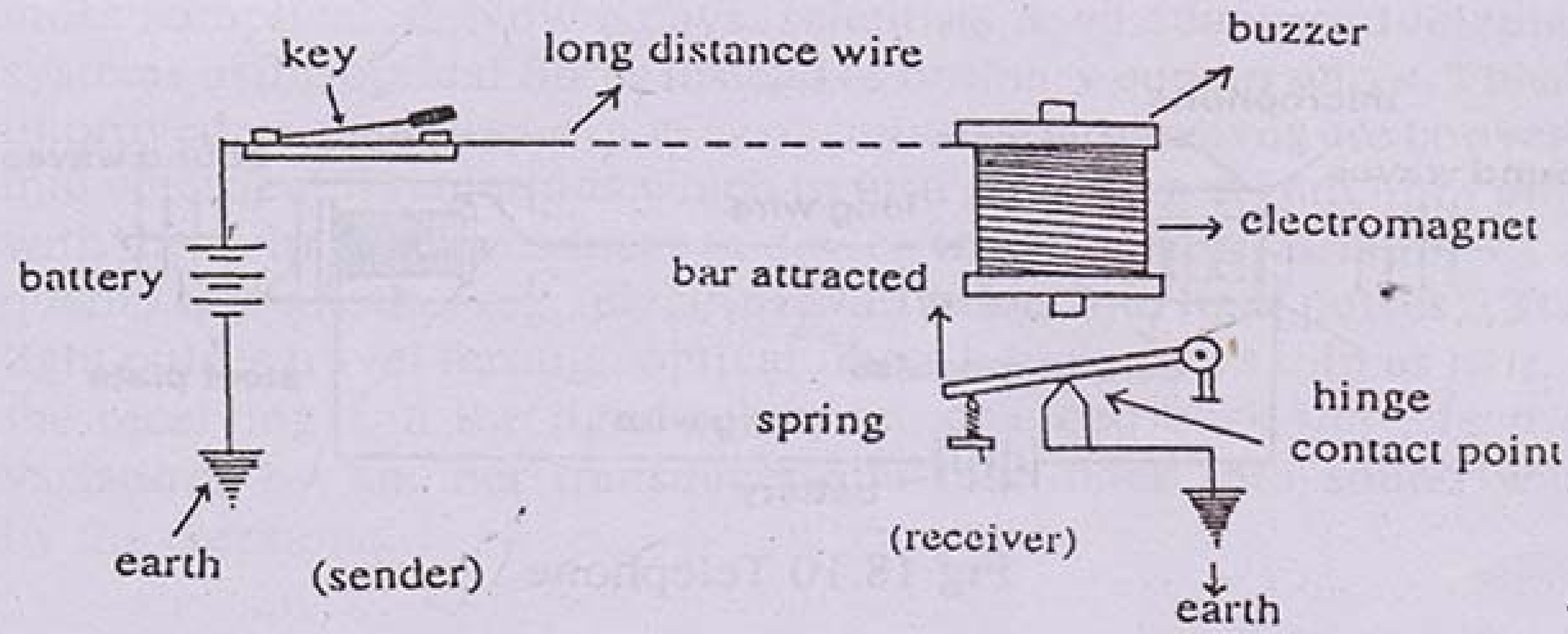


Fig. 18.9 The Telegraph

The buzzer consists of an electromagnet connected through a movable iron bar which is held against a contact point by a light spring. When tapping key is pressed, a current flows through the coil of the electromagnet and the bar is attracted towards the electromegnet. This breaks the circuit between the bar and the contact point. The current stops flowing and the bar is no longer attracted. The spring then pulls the bar back to the contact point and the current again flows. Thus by pressing the tapping key the iron bar vibrates producing a buzzing sound. The interval between two buzzing sounds can be controlled by the interval between pressing the tapping key. A short interval is called a dot and a long interval is called a dash. By using International Morse Code, messages can be sent from one place to another distant place.

In telegraphy a message can be sent and received at the other end of a telegraphy network by experts who know the International Morse Code very well. This mode of communication is also slow and time consuming. Thus there was a need to have such a device by which speech could be transported over a long distance directly from the speaker to the listener. The telephone, which sends sound to a distant place, is the appliance that fulfills this need. This is based on the same principle as that of the telegraph, i.e. sending and receiving messages between two distant places by an electric current flowing through the wires connecting two places.

In the telephone system, invented by Graham bell in 1876, a carbon microphone replaces the tapping key and a magnetic earphone takes the place of the buzzer. A simple circuit diagram is shown in Fig.18.10.

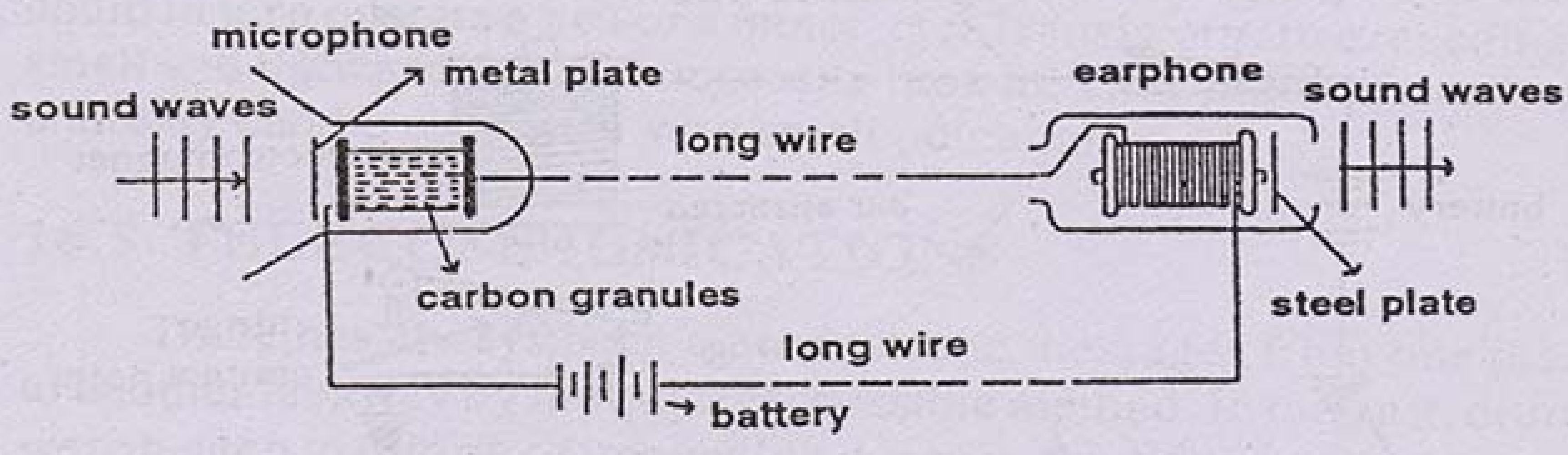


Fig 18.10 Telephone

The carbon microphone has a thin metal plate called a diaphragm suspended in front of a packing of carbon granules. When these granules are in a compressed state the air space or inter-granular distance decreases and so the resistance offered by the layer of carbon granules to the flow of current decreases. More current flows when the granules are in the compressed state than when the distance between the granules is relatively large. If some one speaks in front of a microphone the compression and rarefaction, that constitute a sound wave, cause the diaphragm to vibrate. These vibrations increase or decrease the pressure on the carbon granules very rapidly. Since the granules are part of the circuit the current fluctuates in harmony with the sound waves.

On the other side the magnetic earphone has a thin steel plate suspended in front of an electromagnet. When the current in the circuit fluctuates the field of the electromagnet changes accordingly. The steel plate therefore experiences fluctuating force of attraction and vibrates in the same fashion as the current in the circuit changes. Sound waves are converted into electrical fluctuations by the microphone at one end and the electrical fluctuations are converted into Sound waves by the earphone at the other end.

It may be pointed out that two wires are used for telephones whereas in telegraphy one wire carries the current while the earth connections complete the circuit. The reason is that the earth connections, if also made in a telephone network, would cause so much noise and distortion in the original sound that the message would become unintelligible.

The above presentation is in fact a simple illustration of the basic principle of a telephone. Modern telephone circuits are however, much

more complicated. Now a days, scientists have developed telephone systems using optical fibres instead of ordinary copper wires. This has improved very much the quality of sound. Sound waves are converted into electrical fluctuations which in turn are changed into light pulses with the help of a transducer (a device that converts variation of one quality into another e.g., electrical variations into light pulses). These light pulses travel through optical fibres which are as thin as hair. On the receiving end the light pulses are changed back into electrical variations by another transducer and then back into sound waves by the earphone.

Radio

The wireless telegraph come into being in 1901, when Guglieimo Marconi sent Morse Code messages across the Atlantic Ocean with a spark transmitter. This proved to be of great assistance in the transmission of information from one place to another but could not be used for the transmission of speech or music. In 1906 human voice was transmitted for the first time. Radio broadcasting of speech and music developed rapidly during the 1920's. Radio is now playing an ever increasing role in human life on land, on sea and in the air. It is being used not only for communication of news, entertainment and messages from place to place but also plays an important role in navigation. The basic principle and working of a radio will be outlined below.

We need transmitting station from where information is sent out through the antena into atmosphere. At the other end people have receiving equipment connected to antenna which receives this information.

When we send a parcel to some one it is taken away by a postman who rushes it to its address through postal service. In electronic communication by radio, high frequency radio waves play the role of postman and the information (Parcel) is hidden in these by a process known as modulation. At the receiving end, container recieves the full package of high frequencies radio waves plus the information and the radio circuit seperates the information by a process known as demodulation. Radio transmitters use alternating current of frequencies ranging from 30kHz to 30MHz, called radio frequencies.

The frequency of audible sound varies from 20Hz to 20kHz and sound waves travel with a speed of about 340 metre per second only.

When some one speaks into the microphone at the radio station, sound waves are converted into electrical fluctuations of current coming from the microphone. This is superimposed on the high frequency alternating current passing in the transmitter antenna with the result that the amplitude of the radio waves also fluctuates correspondingly. The sound or message is, therefore, carried through space by these fluctuating radio waves, which are known as modulated carrier waves.

When the modulated carrier waves meet a receiving aerial, which is a piece of metal wire, at some distance from the broadcasting radio station, they generate fluctuating alternating current in it. An earphone is connected to the receiving aerial through a suitable rectifier. The rectifier converts the alternating current generated by the radio waves into direct current. This direct current fluctuates in harmony with the changing amplitude of the radio waves, thus causing the earphone to reproduce the sound.

There are usually a number of radio stations in an area broadcasting their programmes at the same time. However, each broadcasting station produces radio waves of a different frequency. All these waves generate alternating currents of their own frequencies on the receiving aerial of a given radio set. We can select one particular frequency from this mixture by using a *tunning circuit* which responds to the desired frequency, and ignores all other frequencies. The better the quality of the radio and its tunning circuit the less it is affected by the waves from other radio station using similar frequencies.

Television

In a television network images of objects, whether moving or stationary, are transmited along with sound over electromagnetic carrier waves from one place and are received at another distant place. Television technology has now been standardized and finds an increasing use in the realm of entertainment, education, industry and space exploration.

For picture transmission, a TV camera is focused on the object or the scene to be televised. The convex lens of the camera produces an image of the object/scene on a thin photo-sensitive plate, called the mosaic, in the camera tube, the material of the mosaic screen has the ability to emit electrons when light strikes it. Where the light is stronges more electrons are given off making the material positive at

that location. A beam of electrons from an electron gun in the camera tube is made to scan the rear surface of the mosaic screen along successive horizontal lines on it (Fig. 18.11). This is achieved by means of a special magnetic deflection system incorporated in the camera tube. When the beam hits an area with high positive charge few of the negative electrons are repelled. When there is only a little positive charge, more of the electrons are repelled. These electrons are collected and converted into voltage pulses. This is how a picture is converted into voltage pulses, called video signals. Sound is also converted into electric pulses known as *audio signals*. Amplified video and audio signals are used to modulate very high frequency carrier signals before being radiated through an antenna in the form of electromagnetic waves.

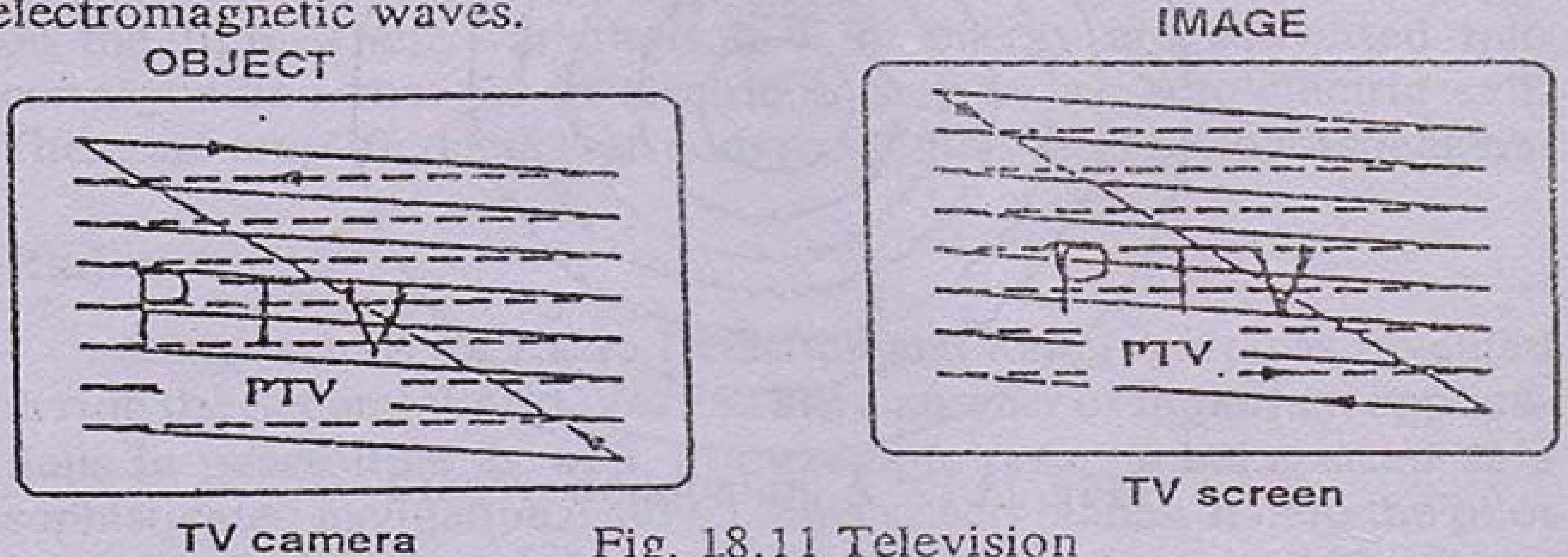


Fig. 18.11 Television

When the modulated electromagnetic waves sent by a TV station strike the antenna of the TV set, they induce fluctuating alternating currents in it. The electronic circuits in the TV set separate the audio and video signals, which are then amplified. The audio signals are sent to the loudspeaker which changes them back into sound waves. Video signals are sent to the electron gun of the picture tube to produce a weak electron beam for darker parts and a stronger beam for brighter parts of the televised picture. The screen of the picture tube is coated with a fluorescent material which emits light when electrons hit it . In this way we can see the picture on the TV set along with the sound.

Sound Recording

All sounds produce vibrations in a material medium as explained in chapter 12. If these vibrations are recorded on some sort of disc or tape, one can reproduce, at any time, the original sound from the impressions made on the disc or tape during the recording process. In

the past gramophone records or discs were very popular for listening to recorded music. The technique of preparing commercial records or discs is as follows.

Sound is converted into fluctuating electric current by a microphone. This current is then amplified and actuates an electromagnetic head with an attached cutting tool. The cutting tool cuts a wavy groove running from the edge to the centre on a metalic plate (Fig. 18.12).

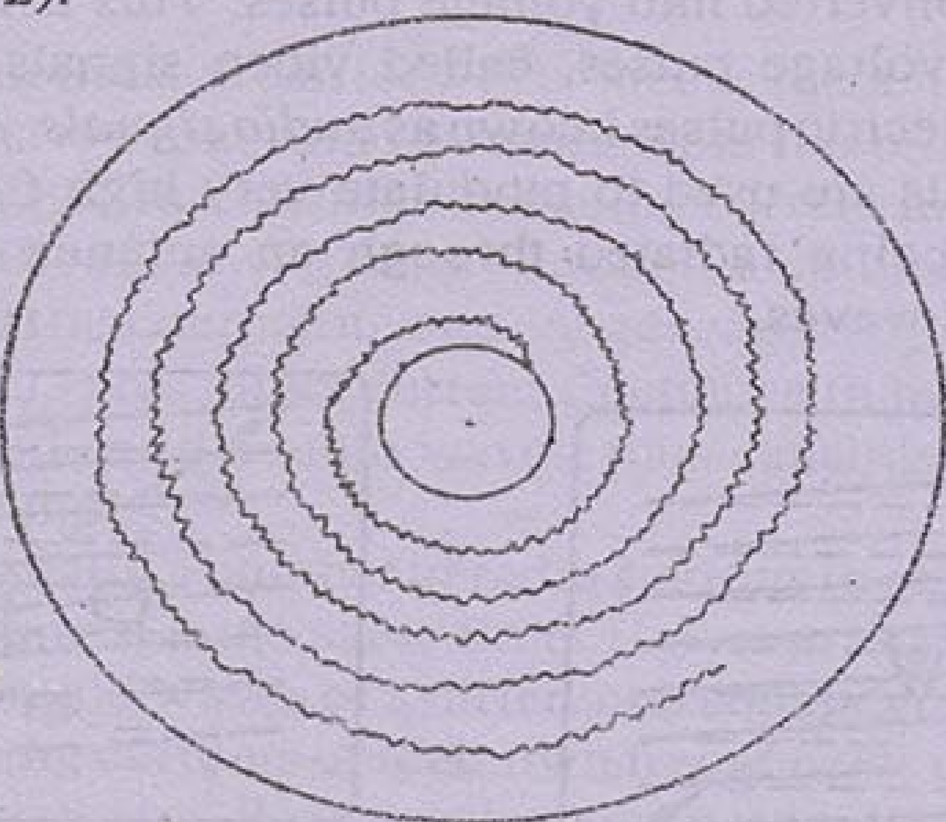


Fig. 18.12 A gramophone record

Sound is reproduced from the record by placing it on the turn table of a gramophone. The turntable revolves with a set speed and a fine needle fixed in a pickup arm is forced to move to and fro laterally by the wavy groove. The vibrations of the needle are converted into a fluctuating electric current, which is amplified and then sent to a loud speaker. The recorded sound is thus reproduced.

The technique of sound recording used in modern tape recorders is much easier. In a tape recorder, when some one speaks in front of a microphone, sound is converted into fluctuating electric current. This current after amplification actuates an electromagnetic recording head. Meanwhile a plastic tape coated with a magnetizable material ferric oxide, chromium oxide is passed in front of the recording head. A continuously varying, magnetic field produced by the recording head magnetises different parts of the moving tape with varying intensity. Sound is thus recorded on the tape in the form of magnetic patterns. To produce sound the recorded tape is passed in front of the playback head which gives rise to a (in proportion to the

magnetism on different parts of the tape) fluctuating electric current. This current is amplified and then passed onto a loud speaker. This is how sound is produced from the recorded tape. In contrast with the gramophone records, the sound recorded on the tape can be wiped out by demagnetization, and the same tape can be used over and over again.

The sound track which accompanies a motion picture is sometimes recorded by electromechanical devices as a small variable density sound track on the edge of the film. The sound waves are converted into a small transparent strip on the side of the film with variations in the film density corresponding to the variations in the voice current. Light that is allowed to shine through the sound track will then have variations corresponding to the voice or music recorded on the film. These variations in light energy are converted into corresponding changes in electric energy by a photo-electric cell. These are amplified and then converted into sound by loud-speakers.

Radar

Radar stands for Radio Detection and Ranging. It was invented during the Second World War and has a number of important applications in peace time as well. For example radar is being used as a sophisticated navigational aid on ships as aeroplanes. It help the pilot in landing a plane even in low visibility due to fog or at night. Similarly it helps the captain of a ship to beware of other ships in the surroundings, icebergs and hidden rocks in the sea. Every modern airport is equiped with a radar system.

Radar consists of a transmitter, a receiver and several idicating devices. The transmitter genertes very high frequency electromagnetic waves (above 600 MHz) which are sent out in any desired direction in a narrow cone shaped beam with the help of a concave antenna. The radar waves travel outward with the velocity of light and are reflected back when they strike a distant object (Fig. 18.13). The reflected wave energy which returns and strikes the radar antenna is amplified in the radar receiver, and these strong signals are fed to the desired indicating devices. These devices measure the time taken by the radar waves to strike the object and come back. By knowing the wave velocity the distance of the object from the radar can be found. The radar waves penetrate fog, haze, smoke and clouds. This makes radar useful for air and surface navigation military purposes as well as tracking satellites

and space vehicles. Other important uses of radar are for air-traffic control, weather observations and storm warnings.

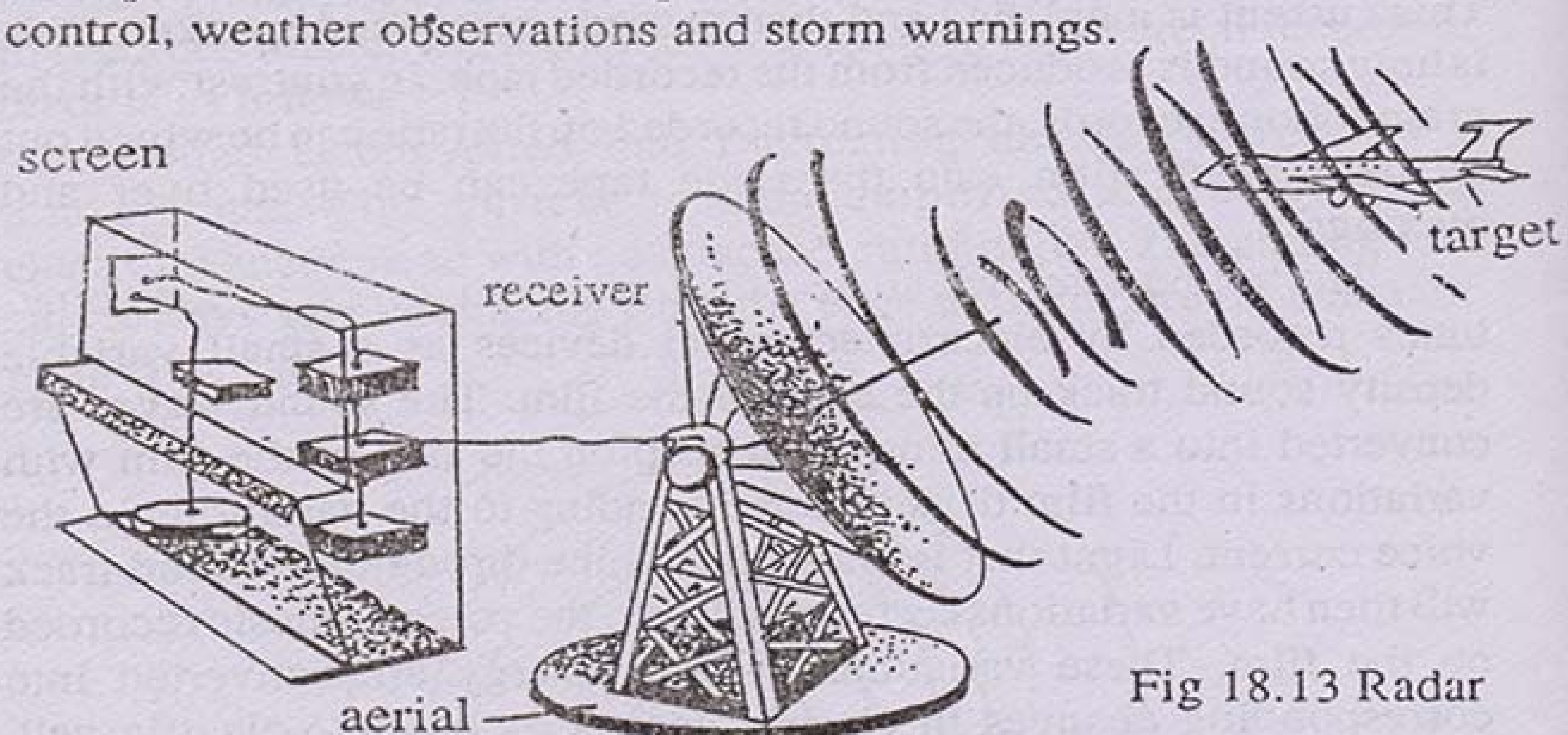


Fig 18.13 Radar

18.6 COMMUNICATION VIA SATELLITE

You may have watched on a Television set a cricket match being played several thousand kilometres away in Australia or England and being shown live in Pakistan. This has been possible due to communication satellites orbiting around the earth (Fig.18.14). For global communication purposes the sattellite launched into an orbit around the earth must be synchronous. It means that they must keep their position static relative to the earth. This is possible when the orbital velocity of the satellite matches the spinning velocity of the earth. The satellites which appear to hover motionless above the same point on earth are called hovering satellites and their orbits are termed geostationary orbits. There are several hundreds of communication and weather satellites in geostationary orbits at various locations around the earth 36000 km above the equator.

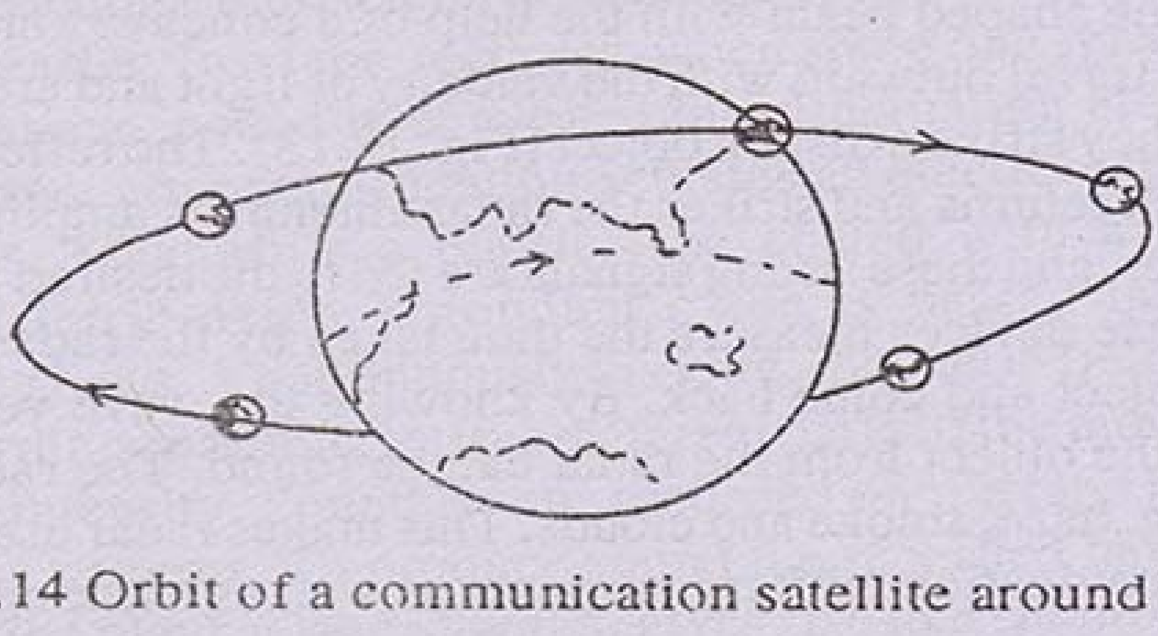


Fig.18.14 Orbit of a communication satellite around the earth.

Modern communication satellites are often powered by solar cells which convert the sun's radiant energy into electricity. Nuclear energy is also utilized to provide power to operate communication satellites. These are equipped with necessary sophisticated electronic equipment and circuitry for the reception and transmission of thousands of telephone conversations simultaneously. They are also designed for TV transmissions to cover the entire world.

High-power, highly-directive, land-based transmitters send wide band microwave signals to the communication satellite above the transmitter. They receive the transmission, amplify it, and re-transmit it to a narrow region on the earth below. In fact three geostationary communication satellites placed in equatorial orbit at 120° from one another can cover practically the whole populated land of the world.

SUMMARY

– Electronics is the branch of physics which deals with the development of electron emitting devices, their utilization and controlling electron flow in electrical circuits designed for various purposes.
– A pure semiconductor material doped with a pentavalent element is called an n-type substance.
– A pure semiconductor material doped, with a trivalent element is called a p-type substance.
– The common boundary of n-type and p-type regions in a semiconductor is called p-n junction diode.
– If the p-type material of a semiconductor diode is at a positive potential and the n-type material is at a negative potential, the diode has a very low electrical resistance, and is said to be forward biased.
– If the p-type material of a semiconductor diode is at a negative potential and the n-type material is at a positive potential, the diode has a very high electrical resistance, and is said to be reverse biased.
– A transistor is a semiconductor which consists of a thin central layer of one type of semiconductor material sandwiched between two relatively thick pieces of the other type.
– Relaying messages, eg. speech, pictures, or in any other form, from one place to another by means of electromagnetic waves is called telecommunication.
– Telegraphy messages are sent and received between two distant places by international Morse Code through an electric current carrying wire connecting the two places.

– The telephone is a device by which two persons at distant places through can directly talk to each other through electric current carrying wires.
– In a radio network the speech or music transmitted through electromegnetic waves from one place can be received at large distances by radio receivers.
– In a television network pictures as well as sound are transmitted through electromagnetic waves from one place and are received at far off places.
– Sound waves are converted into mechanical waves. Which are recorded on gramophone discs. Sound waves can also be converted into pulsating magnetic fields, which are recorded on plastic tapes coated with magnetizable material. The sound recorded in either way can be reproduced whenever needed.
– Radar is a device which uses electromagnetic waves for air and surface navigation by detecting objects in the surroundings of a plane or ship.
– Artificial satellites equipped with necessary electronic devices are made to move around the earth at an altitude of 36,000 km with an orbital velocity equal to the spinning velocity of the earth. These apparently stationary satellites are used to receive signals e.g. speech, conversation, pictures etc, from earth based stations, amplify the signals and re- transmit to the earth to cover longer area transmitted over electromagnetic waves from one place to far off places on the earth surface.

QUESTIONS

18.1 Write answers to the following questions.

(i) What are semiconducting materials? Describe P-type and N-type materials.
(ii) Show that a p-n junction acts like a diode.
(iii) What are transistors? Explain its working.
(iv) Highly pure silicon and germanium crystals are almost insulators, specially at low temperatures. Explain why?
(v) Electric current can easily pass through the p-n junction, if the p-material is at positive potential and n-material at negative potential. Explain.
(vi) A junction diode always allows electric current to pass through it in one direction only. Why?

(vii) What is meant by forward and reverse bias?
(viii) How a message is transmitted by a telegraph?
(ix) How does a television camera work to produce a picture?

18.2 Fill in the blanks

(i) The semiconductors doped with pentavalent elements are called __________ substances.
(ii) A semiconductor diode has a very __________ electric resistance when it is forward biased.
(iii) When a semiconductor diode is reverse biased, electrons and holes will move __________ the p-n, junctions.
(iv) The sound is carried through space by fluctuating radio waves, also known as __________ carrier waves.
(v) The __________ satellites launched into an orbit around the earth must keep their position static relative to the earth.
(vi) Group __________ impurities are called accepter impurities.
(vii) Group v impurities are called __________ impurities.
(viii) Holes are the charge carrier in __________ type material.
(ix) __________ are the charge carrier in N-type material.
(x) PN junction behaves as a __________.

18.3 Given below are a few possible answers to each statement, of which one is correct. Identify the correct one.

(i) The materials in which electric current can flow easily due to their low resistance are called __________.
(a) insulators (b) semiconductors (c) conductors
(ii) The electric resistance of a semiconductor __________ if temperature is increased.
(a) decreases (b) increases (c) does not change
(iii) A p-type substance is formed when a semiconductor crystal is doped with a __________ element.
(a) divalent (b) tetravalent (c) pentavalent (d) trivalent
(iv) The current passing through a __________ is directly proportional to the potential dirference across its ends.
(i) insulator (b) semiconductor (c) conductor
(v) The frequency of radio waves lie in the range __________.
(a) 20 Hzto 20 kHz (b) 30 kHz to 30 MHz
(c) 600 MHz and above.

(vi) __________ geostationary communication satellite placed in equatorial orbit at 120° from one another, can cover the whole populated land of the world.
(a) Three (b) Four (c) Five

(vii) A radio wave is produced by a __________ alternating current
(a) Low frequency (b) high frequency (c) audio frequency

(viii) To produce N-type crystal, germanium may be doped with a substance that is.
(a) divalent (b) trivalent (c) tetravalent (d) pentavalent

(ix) For forward biasing a PN junction, the positive terminal is connected to.
(a) P-type crystal (b) N-type crystal
(c) neither P or N type crystal.

(x) In p-type semiconductor most of current is carried by
(a) electrons (b) protons (c) holes.

18.4 Pick out true and false from the following sentences.

(i) Rubber, glass, mica, ceramics and plastics are good conductor.
(ii) Semiconductors are the elements in the group IV.
(iii) The electric current can hardly pass through the p-n junction if a semiconductor diode is forward biased.
(iv) Sound waves travel in space with the velocity of light.
(v) Holes are the charge carrier in N-type material.

CHAPTER - 19

NUCLEAR PHYSICS

19.1 INTRODUCTION

In this chapter we will deal with the central part of an atom called nucleus which is very small and dense with particles consisting of protons and neutrons. The emission of radiations from the unstable nuclei is described. The processes of fission and fusion which are enormous sources of energy are also discussed.

19.2 THE NUCLEUS

It has been discussed in previous classes that the nucleus of an atom is made up of protons and neutrons. The neutron was discovered by James Chadwick in 1932 and was found to carry no electrical charge. The protons carry a positive charge and therefore, the nucleus of an atom is positively charged. The charge on a proton is $1.6\times10^{-19}\,C$ and opposite to that of an electron. The atom as a whole is electrically neutral, then number of protons in the nucleus is equal to the number of electrons around it. It has been determined that the mass of a proton is nearly equal to the mass of a neutron, which is equal to $1.67\times10^{-27}\,kg$. The protons and the neutrons in a nucleus are collectively referred to as nucleons. The mass of each nucleon is about 1835 times greater than an electron. Thus the mass of an atom is mainly due to the mass of the nucleus, i.e., the mass of its protons and neutrons.

The number of protons in the nucleus of an atom is different in different elements and is called its atomic number. It is denoted by the letter Z. The number of neutrons in the nucleus is denoted by the letter N. The total number of protons and neutrons in a nucleus of all atom is called its atomic mass number and it is denoted by the letter A=(Z+N). An atomic nucleus with a given atomic and mass number is written as ${}_ZX^A$, where X is any element. For example, a uranium nucleus of mass number 235 and Atomic number 92 is written in shorthand notation as ${}_{92}U^{235}$ and a helium nucleus of mass number 4 and atomic number 2 is written as ${}_2He^4$. Here the chemical symbols U

and *He* are the symbols for uranium and helium atoms respectively. For a neutron the symbol $_0n^1$ is used.

Experimental measurements show that the nucleus is very small in size and occupies a spherical region in space. The diameter of a nucleus is of the order of 10^{-14} m, which is ten thousand times smaller than the diameter of an atom, and depends on the atomic mass number. The nucleons are clustered together as incompressible pieces of matter in the spherical region of the nucleus.

19.3 NATURAL RADIOACTIVITY

Elements which have mass number greater than 82 such as Plutonium (Z = 84), Radium (Z = 88). Uranium (Z = 92) etc., are unstable by nature. These elements, all the time, emit different types of powerful radiations. This phenomenon is known as *Radioactivity.* The atoms of these element as a result of emission of these radiations undergo a process of decay and they gradually transform atoms of one element into another. Radioactivity is an irreversible process which continues all the time. As it is an inherent and natural characteristic of radio-active elements to radio activity exhibited by these elements is known as *Natural radio activity.*

Radio-activity was discovered accidentally in 1896 by a French scientist *Henry Becquerel.* He was studying some compounds of uranium which glow in darkness when irradiated by ultra-violet rays. He wanted to know whether these compounds emit x-rays. For this purpose, he performed the following simple experiment.

He wrapped a photographic plate by a black paper in such a way that it was not affected by-sunlight even if placed in direct sunlight for the whole day. He placed some crystals of a uranium salt on this plate. On developing the plate, an image of uranium salt crystals was found on it. Becquerel repeated this experiment by placing a thin glass sheet between the uranium salt and the photo-graphic plate. He found the same result. By similar experiments, Becqueral concluded that uranium emits such radiations which can pass through black paper and can affect photographic plate. For further confirmation of his findings, Becqueral performed several other experiments and from each he concluded that very powerful radiations are emitted from uranium and any special physical or chemical conditions of uranium is not essential for this activity. Emission of invisible radiations from uranium and other substances is known as

Radio-activity and the substances emitting such radiations are called radio active.

19.4 ALPHA, BETA AND GAMMA RAYS

Soon after the discovery of radioactivity, Rutherford and other scientists found out that radiations emitted from radio-active elements are of three types named as Alpha, Beta and Gamma rays. These can be seperated by a simple experiment. A small quantity of a radio active substance. Such as radium is so placed in a cavity in a block of lead that the radiations from radium can come out of the mouth of this cavity (Fig.19.1). A photographic plate is placed at some distance above the lead block so that radiation from radium fall upon it. This appa atus is placed in a vaccum light chamber which is evacuated by a powerful pump. This chamber then is placed between the poles of a strong magnetic field is perpendicular to the plane of paper is directed inwards. Under the action of magnetic field,two of the three types of radiations are deflected. So three separate images are formed on the photographic plate. This shows that radiations coming out of radium are of three types.. One of these are those radiations which bend towards left and give rise to a black spot on the plate close to the centre one. The direction and extent of deflection shows that these radiations consists of positively charged particles which are quite massive. Further experiments show that these particles are nuclei of helium. These radiations are known as Alpha rays.

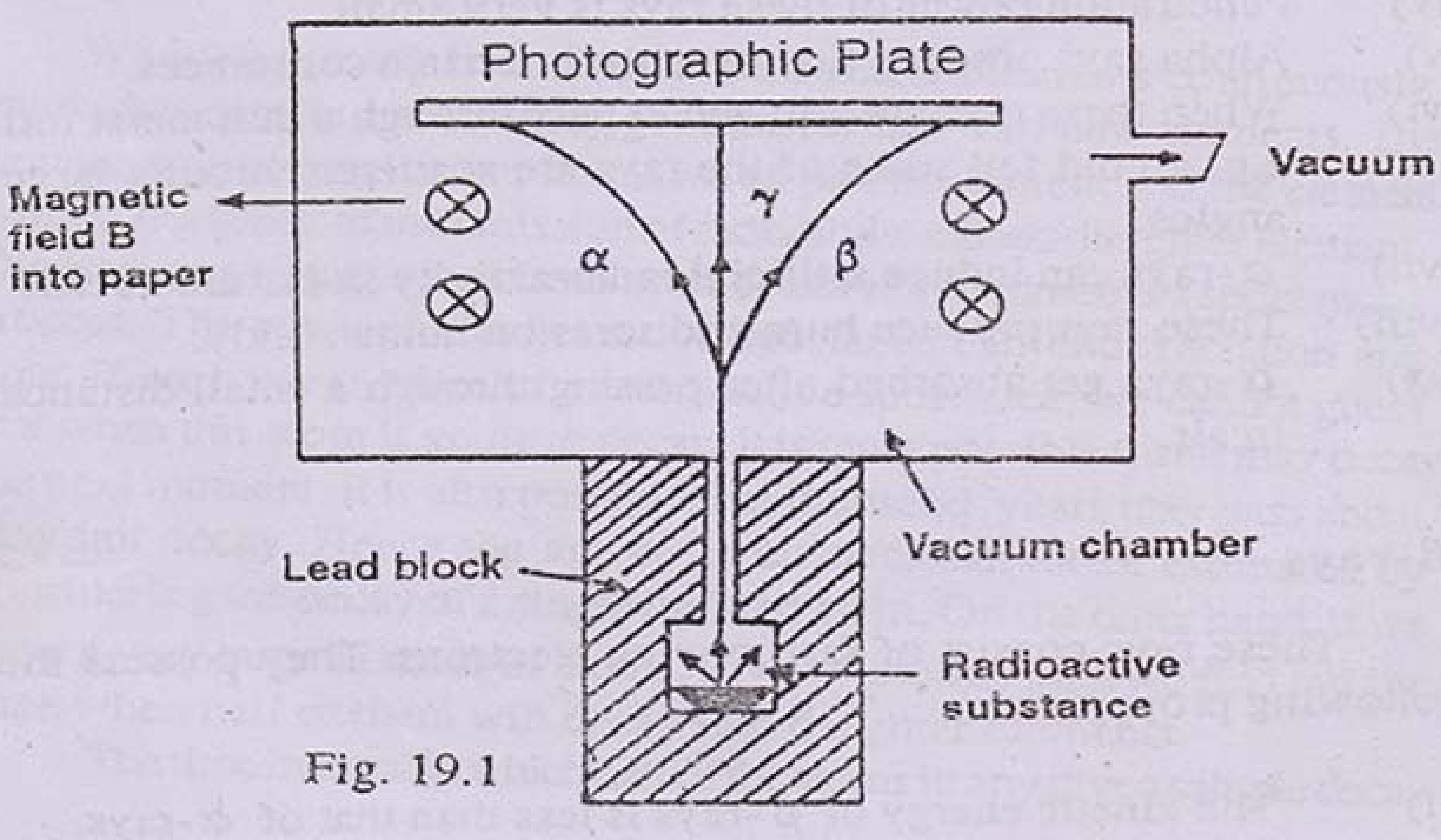


Fig. 19.1

The second type of radiations are those which bend towards right and form a black spot on the plate at a comparatively larger distance from the central spot. This indicates that these radiations consists of negatively charged particles. Later experiments have shown that these particles are actually electrons. These radiations are known as Beta Rays.

The third type of radiations are those which give rise to the black spot at the centre. The fact that these radiations move in the magnetic field without any deflection shows that magnetic field has no effect upon them. This indicates that these rays are neither positively charged particles nor negatively charged particles. Further experiments showed their nature similar to light or X-rays. These radiations are known as Gamma rays.

Properties of Radio-active Rays

α-Rays

Experiments show that α-rays consist of such particle's which are nuclei of helium. These rays possess the following properties.

(i) The mass of each α-particle is nearly four times the mass of hydrogen nucleus.

(ii) The charge on each α-particle is positive and equal to twice the charge on a proton.

(iii) The ionization capability of a α-rays is very large.

(iv) Penetration power of these rays is very small.

(v) Alpha rays produce flourescence in certain substances.

(vi) When these rays are allowed to pass through a thin metal foil e.g. a gold foil some of the rays are scattered through large angles.

(vii) α-rays can induce artificial radio-activity in certain nuclei.

(viii) These rays produce burn and sores on human body.

(ix) α-rays get absorbed after passing through a small distance in air.

β-rays

These rays consist of fast moving electrons. They possess the following properties:

(i) The kinetic energy of β-rays is less than that of α-rays.

(ii) These rays effect the photographic plate.

(iii) These rays produce fluorescence easily, especially in barium platinocynide.

(iv) Due to their small mass, these rays as compared to α-particle, are easily scattered by the nuclei of atoms.

(v) The ionization power of these rays is very small.

(vi) The velocity of β-rays is from $9\times10^7 m/s$, to $27\times10^7 m/s$.

γ-Rays

Since γ-rays are not affected by magnetic field, it is concluded that they do not possess any kind of charge. Experiment show that they are electromagnetic radiation similar to X-rays. Their wave-length is very small but their energy is very high. Some of their other properties are given below.

(i) γ-rays produce feeble fluorescence when incident on a screen coated with barium platinocyanide.

(ii) They eject electron when incident on metals.

(iii) The speed of these rays is equal to that of light.

(iv) Like α-rays, these rays also get absorbed in various materials.

(v) Penetrating power of γ-rays is very large. It is about hundred, times larger than that of β-rays.

19.5 HALF LIFE OF AN ELEMENT

We have already learned that radio-active elements continuously emit radiations, as a result they get transformed into new elements. The element emitting radiations is known as parent element and the element formed as a result of the emission of radiation is called daughter element. The emission of radiations from a radio-active substance is a spontanous process. The atom of a radio-active substance can emit radiation any time. If we have an atom of a radio-active substance, we cannot guess that when this atom is going to decay. It is quite possible that it may decay the next moment. It is also possible that thousand years may pass and it may not decay. Hence the age of an element cannot be estimated by considering the decay of a single isolated atom. On the other hand, if we have a very large number of atoms, say 10^5 atoms, then we find out the time when half of them will decay into daughter elements.

The time interval in which half of the atoms in any given sample decay

into daughter elements is known as the half life of the parent element.

Thus if we start with a sample of 100,000 atoms of a radio active element with a half life of T, then after a lapse of T time, 50,000 atoms of the element will decay into daughter element and the number of remaining parent atoms will be 50,000. After the lapse of another period of time T, the number of atoms of the parent element will reduce to 25,000 and so on. Different radio active material have different half lives which may range from 10^{10} years to a fraction of a second.

Example 19.1

Radium has a half life of 1600 years. How much of 40 gram Radium would be left after 3200 years.

Solution

Number of half lives of radium in 3200 years $= 3200/1600 = 2$ half lives

Amount of radium after first half life $= \frac{1}{2} \times 40g = 20g$

Amount of radium after second half life $= \frac{1}{2} \times 20g = 10g$

19.6 RADIOACTIVE ISOTOPES

Isotopes of an element have the same atomic number but different mass number i .e., they differ only in the number of neutrons. The chemical properties of the isotopes of an element are the same. The nucleus of a carbon atom is usually composed of six protons and six neutron and this carbon atom is represented by $_6C^{12}$. However, there are a few carbon atoms which have eight neutrons in their nuclei and they are represented by $_6C^{14}$. Carbon atoms with mass number 14 are unstable and emit β-radiations. This isotope of carbon is called a radioactive isotope. Thus, radioactive isotopes are unstable isotopes. They may emit α, β or γ radiations.

Table 19.1 Radioactive isotopes and their half-life

Element	Radioactive Isotope	Radiations emitted	Half life
Hydrogen	${}_1H^3$	β	$1.22\times10^{+1}$ yr
Lithium	${}_3Li^8$	β, γ	8.44×10^{-1} s
	${}_3Li^9$	β	1.8×10^{-1} s
Krypton	${}_{36}Kr^{89}$	β, γ	3.16 min
Carbon	${}_6C^{14}$	β	5.73×10^3 yr
Iodine	${}_{53}I^{131}$	β	8.04 days
Polonium	${}_{82}Po^{214}$	α, γ	1.64×10^{-4} yr
Radium	${}_{88}Ra^{226}$	α, γ	1.6×10^3 yr
Indium	${}_{49}In^{115}$	β	4.41×10^{14} yr
Uranium	${}_{92}U^{238}$	α, γ	4.47×10^9 yr
Radon	${}_{86}Rn^{222}$	α, γ	3.83 days

Hydrogen has three isotopes: ${}_1H^1$ with one proton only, deuterium ${}_1D^2$ with one proton and 1 neutron and tritium ${}_1T^3$ with 1 proton and 2 neutrons (Fig.9.2). This isotope is radioactive. In the Table. 19.1, a list of radioactive isotopes with their half life is given.

Many of the isotopes are naturally occuring radioactive isotopes. Radioactive isotopes can also be produced by bombarding subatomic particles onto some elements. For example, a radioactive isotope of cobalt is produced by bombarding it with neutrons.

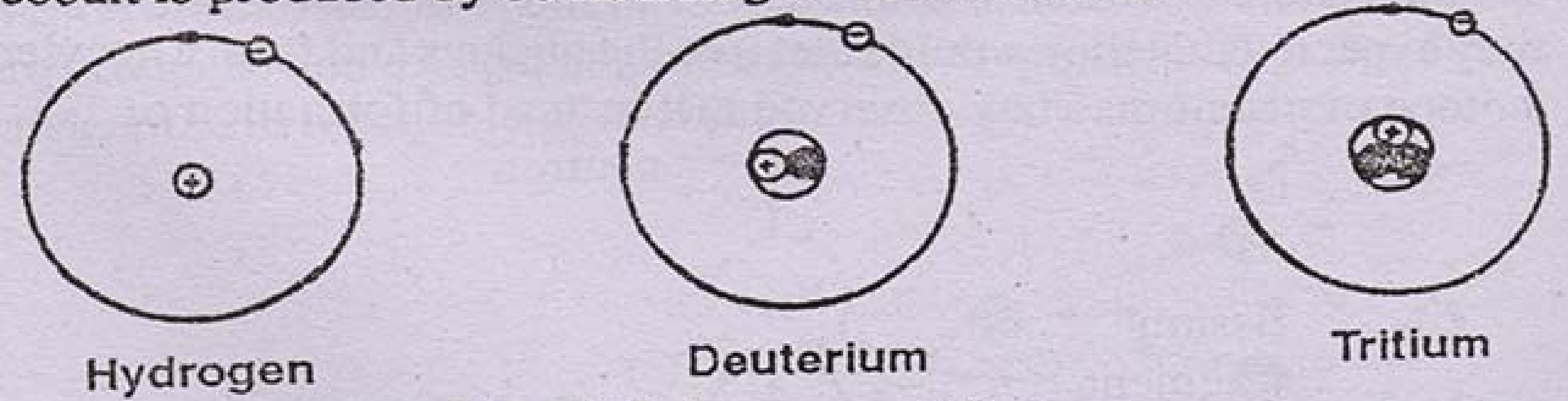


Fig. 19.2 Isotopes of Hydrogen

19.7 EINSTEIN MASS-ENERGY RELATION

In 1905, Einstein proposed the theory of interconvertibility of matter and energy according to the equation.

$$E = mc^2$$

Where E is the energy produced as a result of conversion of m mass of the matter into energy and c is the speed of light. This equation is known as mass energy relation. When Einstein put forward this equation, there was no experimental evidence in its favour and the scientists were reluctant to believe it. But later the enormous release of energy during the fission process could be explained only by the help of this equation.

Example 19.2

In a nuclear reaction 9.0×10^{10} J of energy is released due to conversion of mass into energy. How much mass has been converted to energy?

Solution

Energy, $E = 9.0\times10^{10}\,J$

mass, $m = ?$

Since speed of light, $c = 3\times10^{8}\,m/s$

$$E = mc^2$$

$$m = \frac{E}{c^2} = \frac{9.0\times10^{10}\,J}{(3\times10^{8}\,m/s)^2} = 1\times10^{-6}\,kg$$

19.8 NUCLEAR FISSION

Otto Hahn and Fritz Strassmann in the year 1938 bombarded the uranium nucleus with slow neutrons. Neutrons being uncharged particle they expected that they would enter into the nucleus and form a heavier isotope of uranium. They observed that instead of formation of

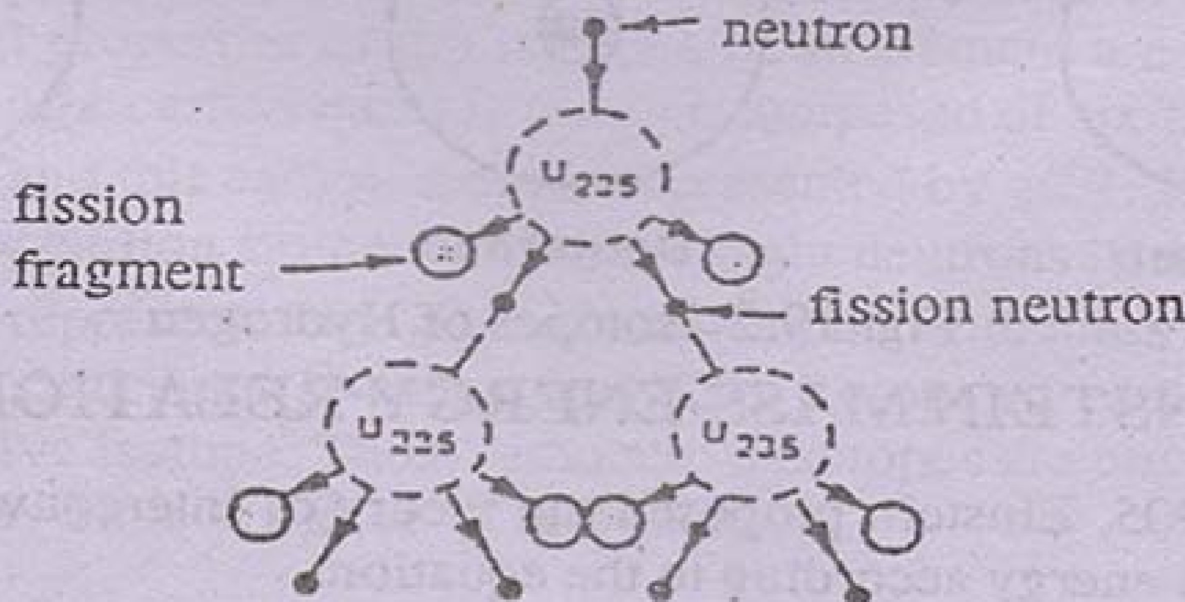


Fig 19.3 A fission process

a heavier nucleus, the nucleus itself broke into two or more fragments with the emission of an enormous amount of energy (Fig. 19.3).

$$\text{Uranium} \xrightarrow{+_0n^1} \text{Uranium (Unstable)} + \text{Fission fragments and energy}$$

The nuclear reaction of the experiment is given as

$$_{92}U^{235} + {}_0n^1 \longrightarrow {}_{56}Ba^{144} + {}_{36}Kr^{90} + 3\,{}_0n^1 + \text{Energy}$$

Where $_{56}B^{144}$ and $_{36}Kr^{90}$ represent the barium and krypton nuclei.

This process was repeated by two German scientists Meitner and Otto Frisch and they, also observed similar results. The splitting of a nucleus into fragments with the emission of energy when bombarded by a neutron is called a fission process. In a fission process the nucleus is split and two to three neutrons are emitted by the splitting of each nucleus. The energy released in a fission process is due to the fact that, when a nucleus splits into fragments, the total mass of the fragments is less than that of the nucleus and the neutron which has caused the nuclear fission. This mass difference according to Einstein's mass energy relation appears as energy.

Chain Reaction

As discussed before, in a fission reaction each nucleus emits about two to three neutrons. These neutrons may collide with the other uranium nuclei and cause fission in them. The nuclei which undergo a fission reaction will emit neutrons. These neutrons will produce further fission in other nuclei. If this process continues, more and more neutrons are produced and a larger amount of energy is released. This is called a nuclear chain reaction. Fig. 19.4 illustrates a nuclear chain reaction.

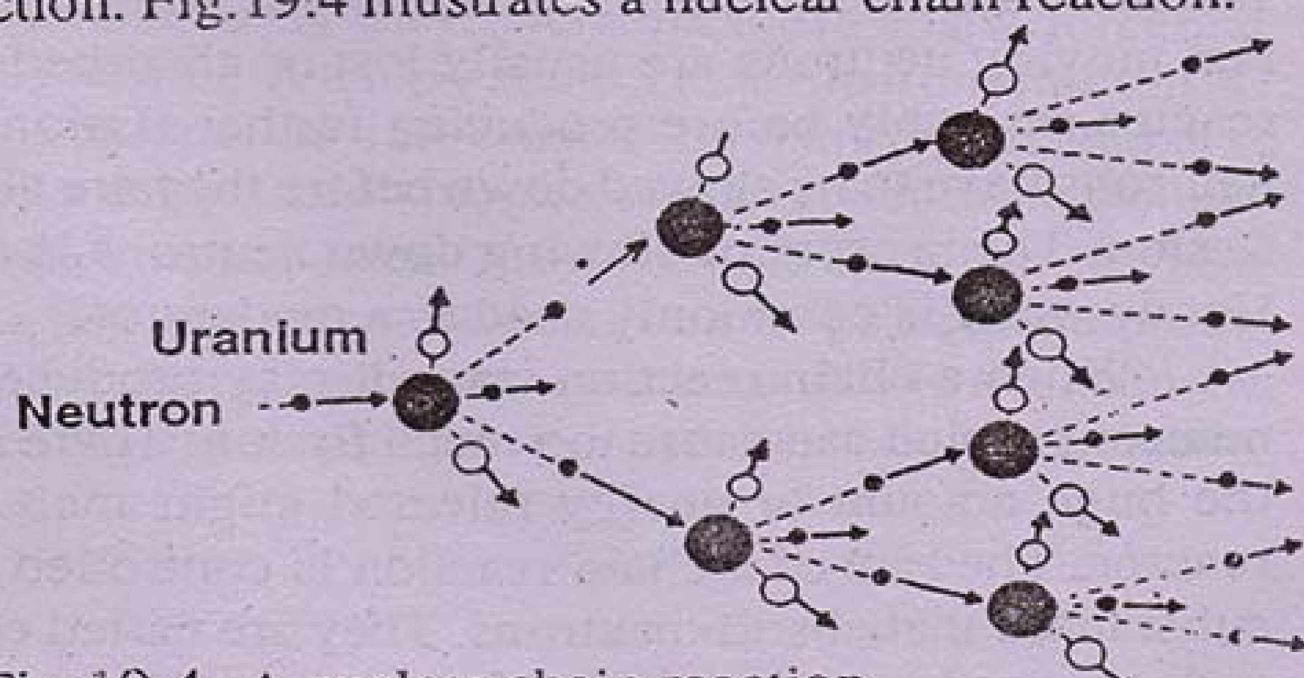


Fig 19.4 A nuclear chain reaction

Nuclear Reactors

As discussed above a nuclear chain reaction releases large amount of energy. This energy is in the form of heat. If the chain reaction is controlled we get a steady outflow of heat. This heat can be used to run a turbine for the generation of electricity. A system used to obtain a controlled amount of heat from nuclear fission is called a nuclear reactor.

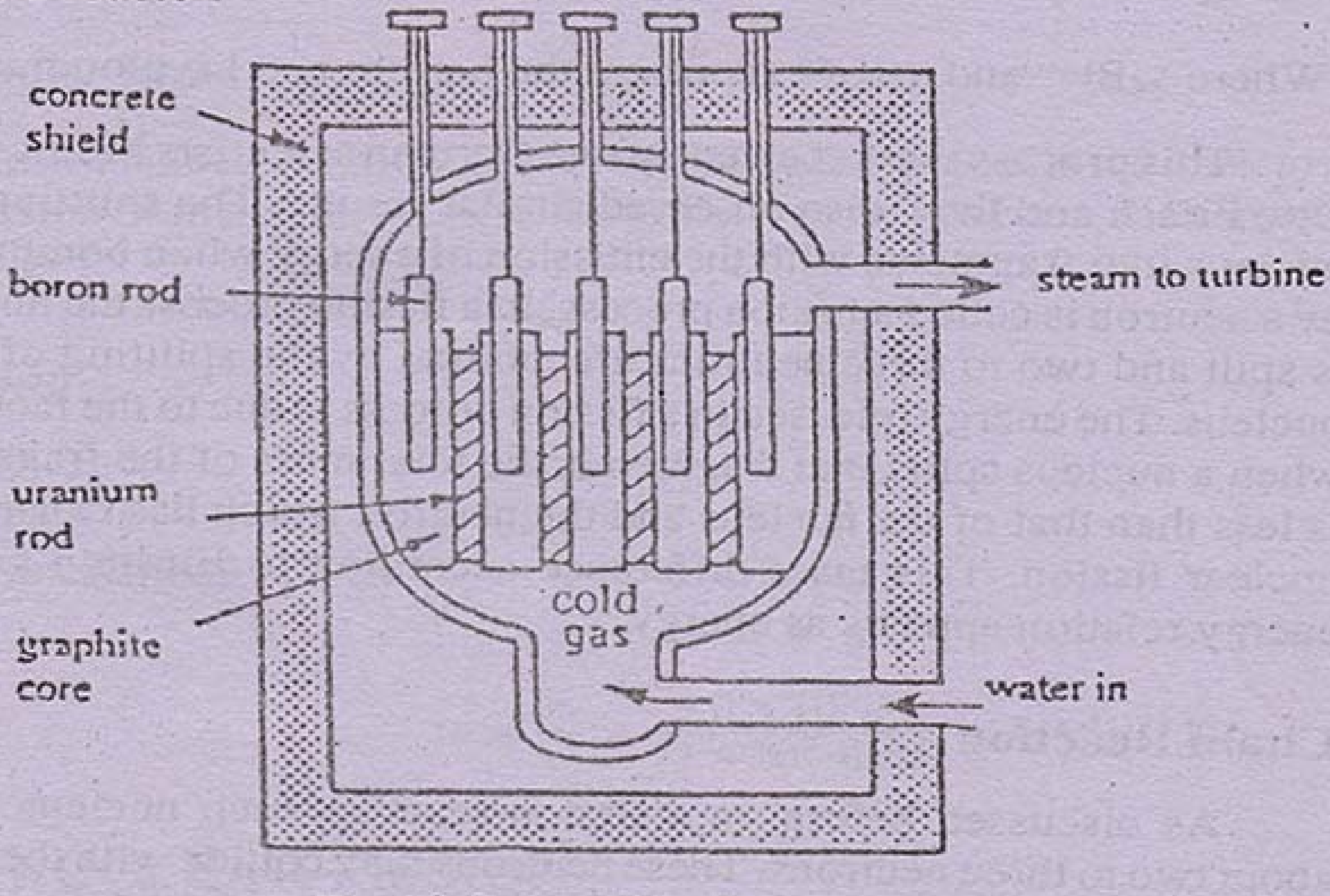


Fig 19.5 Details of a reactor

Fig. 19.5 shows a schematic diagram of a nuclear reactor. The fission material in a reactor is uranium $_{92}U^{235}$. This is called the fuel element. The neutrons released from fission move with high velocities. These fast moving neutrons are usually lost or absorbed somewhere in the reactor assembly before producing further fission. The fast moving neutrons have to be slowed down before they are able to cause further fission. The process of slowing down neutrons is called moderation. Heavy water is commonly used as a moderator.

When a chain reaction starts it may produce large numbers of neutrons which can cause too much fission. There will be danger that the huge amount or energy released might melt down the nuclear reactor. The rate of a chain reaction is controlled by inserting some substances that absorb neutrons. They are called control rods. Boron rods are used as control rods. If too many of the neutrons are absorbed

by the control rods, the chain reaction will stop. To start the chain reaction the control rods are moved out.

The heat produced in a nuclear reactor is carried away by the circulation of carbon dioxide gas or pressurized water inside the core of the reactor. The carbon dioxide gas, or pressurized water, gets extremely hot while passing through the core of the reactor and is then used to produce steam. This steam, can be used to run a power station for the generation of electricity.

19.9 NUCLEAR FUSION

In nuclear fusion light nuclei are brought together to form a relatively heavier nucleus. The energy released in nuclear fusion is larger than that released in fission. When deuterium and tritium muclei are brought together, they form a helium nucleus, with the release of a large amount of energy and a neutron. This process can be represented by the nuclear reaction.

$$_1H^2 + {}_1H^3 \longrightarrow {}_2He^4 + {}_0n^1 + \text{energy}$$

In this process the sum of the masses of the helium nucleus and the neutron is less than the sum of masses of the deuterium and tritium nuclei. This difference in mass according to Einstein's mass energy relation released as energy, given by Δmc^2.

The major difficulty in getting a fusion reaction to take place is that two positive charged nuclei have to be brought together to cause the reaction. For this purpose the two particles must have a huge amount of kinetic energies to overcome the electrostatic repulsion as the two nuclei approach each other. The only way to supply so much energy to the two nuclei, so that they, overcome the repulsive force and fuse together, is by heating. A temperature of more than one million °C is required for this purpose. In a Hydrogen bomb this much temperature is attained by a using fission process to start the fusion reaction.

No fusion reaction has so far been possible in which the heat obtained from the fusion process can be utilized for peaceful purposes. This is due to the fact that scientists have not been able to maintain the high temperature for long enough for the reactions to be maintained.

Energy of the Sun

The largest source of energy in our solar system is the sun. It is strongly believed that the energy output of the sun is due to the fusion process. The temperature inside the sun is about 1.5×10^{6} °C. This temperature is sufficient to cause fusion. The fusion process in the sun has been taking place for the last 5.0×10^{9} years. Hydrogen isotopes serve as the fuel for the fusion reaction. During each second millions of tons of this fuel are consumed. However, the mass of the sun is so great that hydrogen is likely to remain there for billions of years.

19.10 THE ATOMIC BOMB

An atomic bomb is a war weapon. Its assembly is based on the principle that if a fission chain reaction is uncontrolled then the energy released will be enormous. Subcritical masses of uranium (small quantity of uranium which by themselves cannot sustain chain reactions) are placed at the two ends of a hollow tube, such that the two masses can be brought suddenly together by igniting an explosive material. When the two masses combine, they give rise to a critical mass and uncontrollable fission chain reaction starts and the bomb explodes.

In an atomic bomb explosion the heat energy produced by it can destroy a small city. Explosion of an atomic bomb also produces shock waves which move with very high speed and which travel several miles and destroy the buildings which they come across. A nuclear explosion also releases large quantities of radiation. This radiation can be very harmful to living beings. The effect of the radiation remains in the atmosphere for several weeks.

19.11 THE HYDROGEN BOMB

A hydrogen bomb is also a war weapon. The heat energy released in the explosion of a hydrogen bomb is far greater than that released from the explosion of an atom bomb. The principle of the hydrogen bomb assembly is based on the fusion process. The hydrogen bomb needs a tremendous amount of heat for its detonation. That is why a hydrogen bomb is also called a thermonuclear bomb.

The required heat is obtained by a fission bomb. The first hydrogen bomb was exploded on an experimental basis in 1952. This yielded an enormous amount of energy. This was found to be equivalent to 1 million tons of TNT

19.12 THE USES OF RADIO-ISOTOPES

Though all isotopes of an element behave in the same manner chemically, but one distinct advantage of a radio-isotope is that its position can be easily identified by the radiation it emits. This property has made radioisotopes very useful in medicine, agriculture, industry and archaeology. In the following we will describe the use of radio-isotopes in these sectors.

(i) Industry

Estimates of wear and tear in machine tools is extremely important in various industrial processes, especially in moving parts like bearings. If the radioisotope is impregnated into the bearing, the fine bearing filing obtained as a result of wear and tear are washed away by lubricating oil. An examination of the radioactivity of the bearings filing containing oil will reveal the wear of the bearing

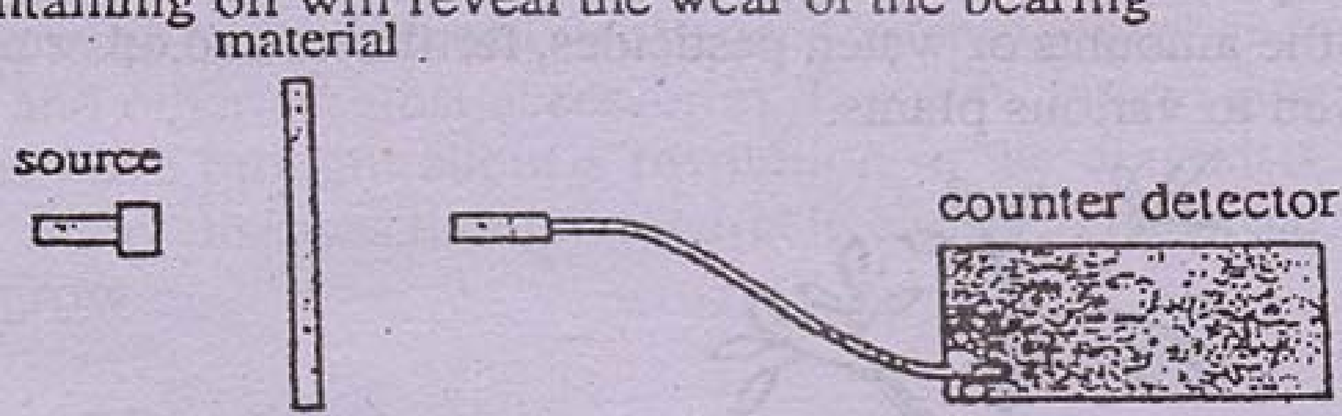


Fig 19.6 Checking the thickness of material using a radioactive source

Radio-isotopes are also used widely to check whether the thickness of a material being produced, is constant or not. As the material passes between the radioactive source and the counter detector (Fig.19.6) any change in thickness causes a change in counting rate in the detector. The irregularity in thickness can thus be exactly pinpointed.

Co-60 emits high energy gamma radiation. It can be used to detect cracks in welded joints.

Radio-isotopes can also be used to detect leakages in pipes. This is done by introducing small quantities of radio isotope into the fluid in the pipe. A radiation detector can be used to check whether the radio isotope is leaking any where in the pipe, thus indicating a fault in the pipe.

(ii) Agriculture

The importance of varieties of seeds, which resist attacks of pests and yield higher production, can not be over-emphasized. Such vari-

eties of seeds for various agricultural commodities have been produced after mutation through radiation. Radiation is also used to kill bacteria and preserve food stuff. One such plant, where fruit, vegetables and other foodstuff is irradiated for preservation has been recently set up in Pakistan by the Atomic Energy Commission.

One very important use of radioisotope is to determine the optimum amount of fertilizers, and other nutrient intake by plants. A very minute amount of radioisotope of the chemical whose absorption by the plant is to be determined is mixed with the ordinary chemical. The whole mixture is then dissolved in water. The plant is then irrigated with this water. A detector placed (Fig. 19.7) near various parts of the plant can identify the location of absorbed chemical, the count rate will determine how much of the chemical has been absorbed in various parts of the plant. In this way the exact location of the absorbed chemical, as well as the quantity of chemical absorbed can be precisely determined. This and similar methods have helped to cut down the amounts of water, pesticides, fertilizers and other nutrients to be given to various plants.

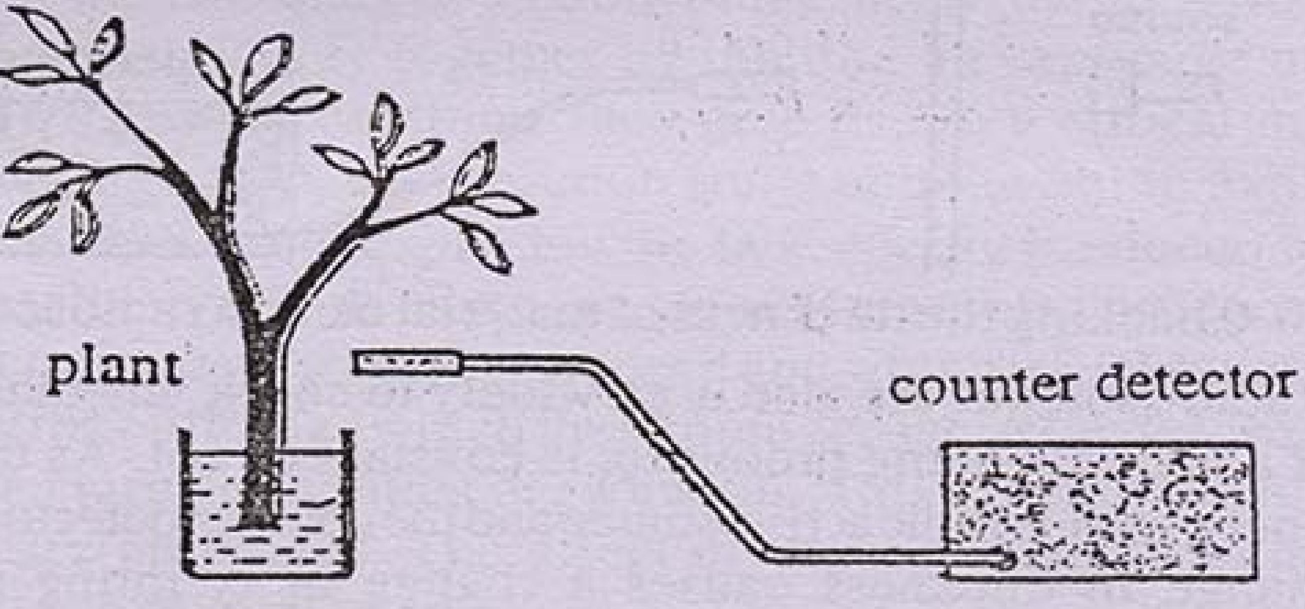


Fig 19.7 - Detecting radiation in a plant

(iii) Medicine

Radioisotopes have helped in understanding the basic working of many of the internal organs and vital metabolic processes. They have also helped in the diagnosis and cure of many complicated diseases. Radioisotope have also played a vital role in determining the effectivity and absorption of medicine in various parts of the bady.

Radio phosphorus-32 and radio-iodine-131 are used as tracer to trace out the path of an element in the body or plant, animal or human being. The position of the radioisotope in the body can be detected

from outside the body by the radiation it emits. This allows some diagnosis about internal organs to be made without surgery. Iodine-131 is used for the study of thyroid glands. Phosphorus-32 is used to locate precisely the position of tumor in the brain. Radiosodium has been extremely useful in tracing the blood circulation in the body. Radio strontium has been found effective in the treatment of internal haemorrhages and wounds. Radio phosphorus has been found effective for treating leukemia. Radiation emitted by it destroys the excess production of white blood corpuscles.

Certain types of cancers also respond to treatment by radiation emitted by radioisotopes. It has been found that certain type of cancerous cells absorb radioisotope radiation preferentially. This helps in locating the cancerous portions of the body precisely, which can then be treated by bombardment by radiation from radioisotopes. Radio cobalt-60 has been widely used to treat cancerous tumors inside the body.

Radiation in low dosages can also be used for sterilizing bandages, instruments and other surgical accessories. In the twentieth century, radioisotopes have brought about a revolution in the diagnosis and treatment of many diseases thought incurable earlier. Their role will continue to grow.

Generation of Electricity

The heat energy released in a nuclear reactor is used for generating electricity. The method used for this purpose is shown in Fig. 19.8

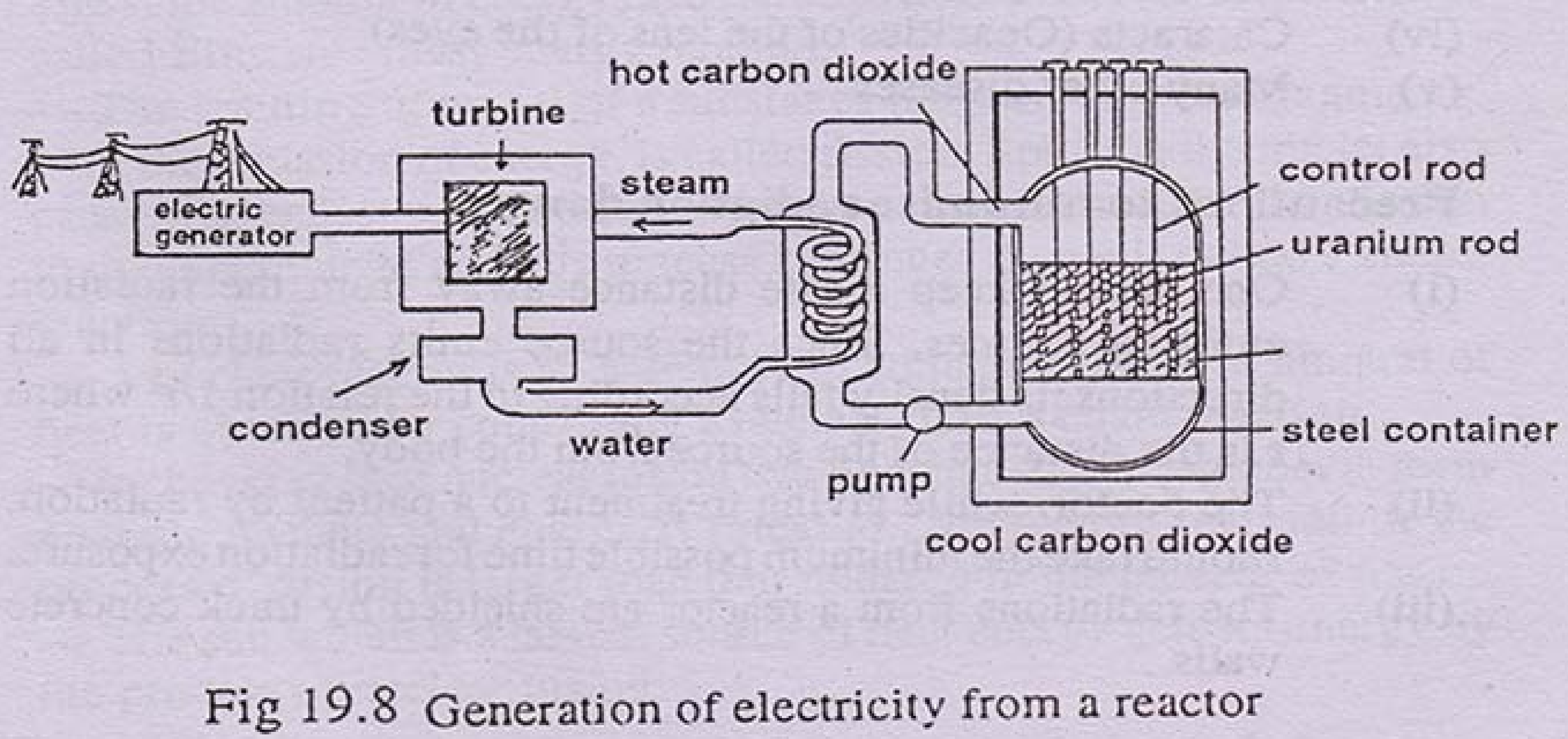


Fig 19.8 Generation of electricity from a reactor

The heat energy produced in a nuclear reactor is carried away by circulating carbon dioxide gas or pressurized water around the reactor core. This hot fluid boils water. The steam produced by boiling water is used to drive the turbine of the electric generator for producing electricity. In this way nuclear fission energy is transformed into electrical energy.

19.13 RADIATION HAZARDS

In contrast to the use of radiation in medicine nuclear radiations can be very dangerous to the human body. They can damage our body cells. Excessive radiations may cause cancer or incurable radiation sickness to a human being.

The destruction of the cell is caused by the ionizing properties of the radiations. The amount of ionization produced depends upon the intensity and energy of radiation. Energetic radiations will penetrate deep into the body and they can destroy the vital body cells. High intensity radiations can destroy large number of vital body cells. The destruction of the body cells also depends upon the exposure time to the radiation. A larger number of body cells will be destroyed if it is irradiated for a longer time. A body, if strongly irradiated, may suffer the following diseases.

(i) Anaemia (a decrease in red blood corpuscles).
(ii) Leukemia (an increase in the white blood corpuscles, also called the blood cancer)
(iii) Malignant tumors
(iv) Cataracts (Opacities of the lens of the eyes)
(v) Many other diseases

Precautions to minimise radiation danger

(i) One should keep a safe distance away from the radiation emitting sources. Since the source emits radiations in all directions its density falls according to the relation $1/r^2$ where r is the distance of the source from the body.
(ii) The doctor, while giving treatment to a patient by radiation, should take the minimum possible time for radiation exposure.
(iii) The radiations from a reactor are shielded by thick concrete walls.

(iv) In a laboratory, where radioactive materials are handled, the radioactive substance is covered in a lead box with a lid made of lead.Lead is a high density material and therefore, stops radiations falling upon it. The experiments are performed in separate rooms, and the students are instructed to handle the apparatus carefully.

SUMMARY

— The nucleus of an atom is composed of neutrons and protons. The neutrons and protons in a nucleus are both termed as nucleons.

— The radiations emitted from a radioactive substance are α, β and γ-rays.

— The α-rays are moving α-particles. Each particle consists of two protons and two neutrons. The α-particle is the nucleus of a helium atom and it carries positive 2 charges.

— The β-rays are fast moving electrons. They carry negative charges.

— γ-rays are very energetic electromagnetic waves. They are emitted when α or β emission leaves the nucleus in an excited state.

— The time during which half the atoms of a radioactive substance decay is called the half-life of that radioactive substance.

— The isotopes of elements have the same atomic number but different mass numbers. Some isotopes are radioactive and they are called radioactive isotopes.

— When mass is converted in energy it releases energy equal to mc^2 The same amount of energy is required to convert it into mass. This is called Einsten's mass energy relation.

— The splitting of nuclei of a substance into two or more fragments, with the emission of energy, is called fission. Splitting the nuclei also causes the emission of neutrons. If these neutrons cause further fission in the other nuclei and this process continues then it is called a chain-reaction.

— A nuclear reactor is an assembly in which a controlled amount of heat is obtained by controlling a nuclear fission chain reaction.

— In the process of nuclear fusion two light nuclei join together to form a single nucleus and energy is released. The energy released in the process of fusion is larger than that released in the process of fission.

— The sun which is a largest source of heat energy gets its energy by the process of nuclear fusion.

— Atomic and hydrogen bombs are war weapons. In an atomic bomb an uncontrolled fission chain reaction occurs and in a hydrogen bomb the process of fusion occurs. These weapons are disastrous for mankind.
— Nuclear energy and the radiations emitted from a nucleus have many uses. For example, heat energy obtained from a nuclear reactor is converted into electricity. The emitted radiations are used in agriculture, medicine and industry.
— Excessive exposure to radiation is dangerous to health. Safety measures are taken to protect from these radiations.

QUESTIONS

19.1 Write answers to the questions given below:

(i) Describe the particles found in the nucleus of an atom. Explain mass number and atomic number.
(ii) What is radio activity? What are alpha, beta and γ-rays? Describe their properties.
(iii) What are radio isotopes?
(iv) What is meant by half-life of a radio active substance?
(v) Describe fission process in some detail.
(vi) What is fusion process?
(vii) Write a note on the usefulness of radio isotopes of elements in medicine, agriculture and industry.
(viii) How can radiation from radioactive substances harm us? How can we protect against there radiotion hazards.

19.2 Fill in the blanks.

(i) The size of the atom is __________ m, where as the size of its __________ is __________ 10^{-14}m.
(ii) Every α-particle carries __________ charge which is double the charge of __________.
(iii) The velocity of β-rays vary from __________ to __________.
(iv) __________ is also called electron.
(v) __________ rays are not deflected by an electric or magnetic field.

(vi) In __________ reaction two lighter nuclie are combined to form a heavy nucleus.

(vii) A reaction in which __________ nucleus splits into fragments is called __________.

(viii) An atom of radon ${}_{86}R_n^{222}$ which contains X protons and Y neutrons, where x = _________ and y = _________.

(ix) Nucleus of _________ is called alpha particle.

(x) The time in which half of the atoms of an element decay is called __________ of an element.

19.3 Given below are a few possible answers to each statement. Identify the correct one.

(i) The number of protons in the nucleus is called _________.
(a) avogadro number (b) atomic number
(c) mass number (d) nucleons number.

(ii). The emission of rays from the nucleus is called _________.
(a) chemical process (b) atomic process
(c) radio-activity (d) atomic dispersion.

(iii) α-rays are found to be in __________.
(a) electromagnatic waves (b) electrons
(c) fastly moving neutron (d) fastly moving helium nucleus.

(iv) An element whose atoms have same atomic number but different mass number are called __________.
(a) molecule (b) secondary element, (c) isotope.

(v) Which of the following is more penetrating.
(a) α-rays (b) β-rays (c) γ-rays

19.4 Pick out true and false from the following sentences.

(i) The nucleus of ${}_{92}U^{238}$ has equal number of protons and neutrons.

(ii) The emission of rays from radium is an example of atomic fission.

(iii) β-particle is a nucleus of hydrogen

(iv) γ-rays carry positive charge

(v) The mass of helium nucleus is equal to the sum of the mass of two protons and two neutrons.

(vi) The splitting of nucleus of the atom is called fission.

(vii) The fusion process continues in the sun and other stars.

PROBLEMS

19.1 The nucleus of nitrogen contains 9 neutrons. Find its charge number if its mass number is 16,

(7)

19.2 The mass of ${}_6C^{12}$ nucleus in found to be 0.164×10^{-27} kg less than its constituents. Calculate the energy released.

(1.476×10^{-11} J)

19.3 How much energy will be released when 15gm of mass is completely transformed to energy.

(1.35×10^{15} J)

19.4 Radium has a half life of 1600 years. How much of 60gm radium would be left after 4800 years.

(7.5 gm)

19.5 Half life of radon is 3.82 days. How much of a 100 gram sample of radon would be left after 764 days.

(25gm)